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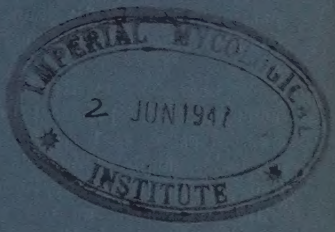


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*MATHEMATICS*

# ON FREE TOPOLOGICAL GROUPS

By A. MARKOFF

*(Communicated by S. N. Bernstein, Member of the Academy, 15. II. 1941)*

1. Theorem 1. *Let $R$ be a completely regular space* (1). *There exists a topological group $F$ with the following properties:*

$F_1$. *$R$ is a subspace of $F$.*

$F_2$. *$R$ generates $F$ algebraically.*

$F_3$. *Given a continuous mapping $\varphi$ of $R$ into any topological group $G$, there exists a continuous homomorphism $\Phi$ of $F$ into $G$ such that $\Phi x = \varphi x$ for every point $x$ of $R$.*

Theorem 2. *If $R$ is a completely regular space, then the topological group with the properties $F_1, F_2, F_3$ is unique up to a topological isomorphism* mapping each point of $R$ on itself.*

Definition 1. The unique topological group with the properties $F_1, F_2, F_3$ will be called the *free topological group of the space $R$.*

2. Theorem 3. *Every completely regular space constitutes an algebraically free system of generators of its free topological group.*

3. Theorem 4. *Every completely regular space is closed in its free topological group.*

Corollary 1. *The free topological group of a completely regular but non-normal space* (1) *is non-normal as a space.*

Corollary 2. *There exist non-normal topological groups.*

4. Definition 2. *$F$* is a *free topological group*, if there exists a completely regular space such that $F$ is its free topological group.

Theorem 5. *Every topological group is topologically isomorphic to a topological factor-group** of a free topological group.*

5. Definition 3. Let $X$ be an arbitrary set. Every finite sequence $\{\{x_i, \varepsilon_i\}\}_{i=1}^{m}$ of pairs $\{x_i, \varepsilon_i\}$ with $x_i \in X$, $\varepsilon_i = \pm 1$ $(i = 1, \ldots, m)$ will be called an *abstract relation* in $X$.

Definition 4. Let $R$ be a completely regular space; $\gamma$, a set of abstract relations in $R$; $\varphi$, a continuous mapping of $R$ into a topological group $G$. We say that $\varphi$ is a *realization of $\gamma$ in $G$*, if the following two conditions are satisfied:

$R_1$. $\prod_{i=1}^{m} (\varphi x_i)^{\varepsilon_i} = 1_G$ as soon as $\{\{x_i, \varepsilon_i\}\}_{i=1}^{m} \in \gamma$;

$R_2$. $\varphi R$ generates $G$ algebraically.

Here $1_G$ denotes the unit element of $G$.

* Our term «topological isomorphism» of topological groups is equivalent to Pontrjagin's «isomorphism» of topological groups (2).

** Our term «topological factor-group» is equivalent to Pontrjagin's «factor-group»(3).

Theorem 6. *Let $\gamma$ be a set of abstract relations in a completely regular space $R$; $\varphi$, a realization of $\gamma$ in a topological group $G$; $\chi$, a continuous homomorphism of $G$ on to a topological group $H$. Then the product $\chi\varphi$ is a realization of $\gamma$ in $H$.*

Definition 5. Let $\gamma$ be a set of abstract relations in a completely regular space $R$; $\varphi$ and $\psi$, realizations of $\gamma$ in topological groups $G$ and $H$, respectively. We say that $\varphi$ *induces* (*is equivalent* to) $\psi$, if there exists a continuous homomorphism (a topological isomorphism) $\chi$ of $G$ on to $H$ such that $\psi = \chi\varphi$.

Definition 6. $\varphi$ is a *realization* of the set $\gamma$ of abstract relations, if there exists a topological group $G$ such that $\varphi$ is a realization of $\gamma$ in $G$.

Definition 7. $\varphi$ is a *primary realization* of $\gamma$, if $\varphi$ is a realization of $\gamma$ inducing every realization of $\gamma$.

Theorem 7. *Every set of abstract relations in a completely regular space possesses a primary realization.*

Theorem 8. *The primary realization of a set of abstract relations in a completely regular space is unique up to equivalence.*

Definition 8. Let $\gamma$ be a set of abstract relations in a completely regular space; $G$, a topological group. We say that $G$ is the *group determined by* $\gamma$, if there exists a primary realization of $\gamma$ in $G$.

Corollary 3. *Every set of abstract relations in a completely regular space determines a topological group.*

Corollary 4. *The topological group determined by a set of abstract relations in a completely regular space is unique up to a topological isomorphism.*

6. The detailed proofs of all these results will be published elsewhere. Here we shall give only a brief sketch of the proof of Theorem 1, since this proof may be considered as a key to all others above formulated results.

This proof is based on the notions of a «norm» and a «multinorm» in a group, defined as follows.

Definition 9. Let $G$ be a group; $N$, a real function in $G$. We say that $N$ is a *norm* in $G$, if the following conditions are satisfied:

$N_1$. $N(1_G) = 0$.

$N_2$. $N(xy^{-1}) \leqq N(x) + N(y)$.

Definition 10. Let $\mathfrak{M}$ be a set of norms in a group $G$. We say that $\mathfrak{M}$ is a *multinorm* in $G$, if the following conditions are satisfied:

$M_1$. $N_1 \in \mathfrak{M}$ and $N_2 \in \mathfrak{M}$ implies $N_1 + N_2 \in \mathfrak{M}$.

$M_2$. $N \in \mathfrak{M}$ and $a \in G$ implies $N_{[a]} \in \mathfrak{M}$.

$M_3$. If $x$ is an element of $G$, such that $N(x) = 0$ for every $N$ belonging to $\mathfrak{M}$, then $x = 1_G$.

Here $N_{[a]}$ denotes the function defined by the equality

$$N_{[a]}(x) = N(a^{-1}xa) \qquad (x \in G).$$

The following theorem can be easily proved with the aid of Theorem 10 of the above-mentioned book of Pontrjagin.

Theorem 9. *Let $\mathfrak{M}$ be a multinorm in a group $G$. For every norm $N$ belonging to $\mathfrak{M}$ let us define the subset $U_N$ of $G$ by the equality*

$$U_N = \underset{x}{\mathcal{E}}\,(N(x) < 1). \tag{1}$$

*Then there exists a unique topologization of the group $G$ such that the totality of all the sets $U_N$ becomes a complete set of neighbourhoods of the unit element.*

Making now use of a method due to Kakutani (4), we can prove the

Theorem 10. *Given a topological group $G$, there exists in $G$ a multinorm $\mathfrak{M}$ such that the totality of all the sets $U_N (N \in \mathfrak{M})$ defined by the equality* (1) *constitutes a complete set of neighbourhoods of the unit element in $G$.*

Theorem 1 can now be proved according to the following scheme. We construct the algebraical free group $F$ with the set $R$ as its free basis. Let $\mathfrak{M}$ denote the totality of all the norms $N$ in $F$ satisfying the following continuity condition: whatever be a finite system $\{\varepsilon_i\}_{i=1}^m$ $(\varepsilon_i = \pm 1,\ i = 1, \ldots, m)$, the expression $N\left(\prod_{i=1}^{m} x_i^{\varepsilon_i}\right)$ is a continuous function of the system of $m$ points $x_1, \ldots, x_m$ of $R$.

We prove that $\mathfrak{M}$ is a multinorm in $F$. Here the proof of $M_3$ presents the only difficult point. According to Theorem 9, $\mathfrak{M}$ yields a topology in $F$; $F$ becomes now a topological group. It remains to prove that $F$ satisfies conditions $F_1, F_2, F_3$. This can be done with the aid of Theorem 10.

7. In an analogous manner applying the results of Hyers ([5]) instead of Theorems 9 and 10, one can prove the following theorems.

Theorem 11. *Let $R$ be a completely regular space. Then there exists a linear topological locally convex space ([6]) $L$ with the following properties:*

$L_1$. *$R$ is a subspace of $L$.*

$L_2$. *Every element of $L$ can be represented in the form $\sum_{i=1}^{m} \lambda_i x_i$, where the $\lambda_i$ are real numbers and $x_i \in R\ (i = 1, \ldots, m)$.*

$L_3$. *For every continuous mapping $\varphi$ of the space $R$ in any linear topological locally convex space $M$ there exists a continuous linear and homogenous mapping $\Phi$ of $L$ into $M$ such that $\Phi x = \varphi x$ for every point $x$ of $R$.*

Theorem 12. *If $R$ is a completely regular space, then the linear topological locally convex space $L$ with the properties $L_1, L_2, L_3$ is unique up to a topological isomorphism, mapping each point of $x$ on itself.*

These results enable us to introduce the notion of the «absolute locally convex bearer» of a completely regular space. We state namely

Definition 11. The unique linear topological locally convex space with the properties $L_1, L_2, L_3$ will be called the *absolute locally convex bearer* of $R$.

We can prove the following results analogous to Theorems 3, 4 and Corollaries 1,2.

Theorem 13. *The points of a completely regular space are linearly independent in the absolute locally convex bearer of this space.*

Theorem 14. *Every completely regular space is closed in its absolute locally convex bearer.*

Corollary 5. *The absolute locally convex bearer of a completely regular but non-normal space is a non-normal space.*

Corollary 6. *There exist non-normal linear topological locally convex spaces.*

A theorem analogous to Theorem 5 can also be proved.

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*MATHÉMATIQUES*

# SUR LA STABILITÉ DANS LE PROBLÈME INVERSE DE LA THÉORIE DU POTENTIEL

Par J. RAPOPORT

*(Présenté par S. L. Soboleff, de l'Académie, le 10. II. 1941)*

Dans le présent article nous indiquons un critère de stabilité dans le problème inverse de la théorie du potentiel, critère contenu dans le théorème suivant ([1], [2]).

Théorème. *Soit $V$ une fonction harmonique définie dans l'espace $xyz$ et qui est le potentiel extérieur d'une masse $M$ de densité constante $\mu$, remplissant un certain domaine $G$, étoilé par rapport à l'origine des coordonnées $O$. Soit ensuite $\sigma^*$ une surface unie, étoilée par rapport au point $O$ et contenant tous les points singuliers de la fonction $V$, $V^*$ — le potentiel extérieur d'une masse $M^*$ de densité constante $\mu$, qui remplit un domaine $G^*$, intérieur à $\sigma^*$. Désignons par $T$ la différence entre les volumes des domaines $G+G^*$ et $G \cdot G^*$. Alors*

$$T < C \iint_{\sigma^*} \left| \frac{\partial W}{\partial n} \right| d\sigma$$

*où*

$$W = \frac{1}{\mu} \left[ 2(V - V^*) - r \frac{\partial}{\partial r} (V - V^*) \right]$$

*$n$ étant la normale extérieure à la surface $\sigma^*$, $C$ — une constante qui ne dépend que de la forme de la surface $\sigma^*$.*

Ainsi les domaines $G$ et $G^*$ seront en moyenne aussi proches que l'on veut, si les fonctions harmoniques $V$ et $V^*$ sont en moyenne assez proches, ainsi que leurs dérivées partielles des deux premiers ordres sur la surface $\sigma^*$. Passons maintenant à la démonstration du théorème énoncé.

Démonstration. Considérons une surface fermée et unie arbitraire $\Sigma$, qui contient à son intérieur le domaine $G+G^*$ et une fonction arbitraire $U$ harmonique sur $\Sigma$ et à l'intérieur de $\Sigma$. Considérons l'intégrale

$$I = -\frac{1}{4\pi} \iint_{\Sigma} \left( U \frac{\partial W}{\partial n} - W \frac{\partial U}{\partial n} \right) d\sigma \qquad (1)$$

où $n$ est la normale extérieure à $\Sigma$. Désignons par $\rho$ la densité d'une couche simple sur la surface $\sigma^*$ dont le potentiel extérieur est $W$. En posant dans (1) $W = \iint_{\sigma^*} \frac{\rho d\sigma}{R}$, nous trouvons:

$$I = -\frac{1}{4\pi} \iint_{\sigma^*} \rho \left[ \iint_{\Sigma} \left( U \frac{\partial R^{-1}}{\partial n} - R^{-1} \frac{\partial U}{\partial n} \right) d\sigma \right] d\sigma = \iint_{\sigma^*} \rho U \, d\sigma. \qquad (2)$$

Introduisons ensuite les notations: $S$ et $s$ sont les surfaces qui bornent les domaines $G+G^*$ et $G\cdot G^*$, $\Delta$ — un domaine compris entre les surfaces $S$ et $s$. Alors on a sur $\Sigma$

$$V-V^*=\iiint_{\Delta}\frac{\nu\mu\,d\tau}{R}, \tag{3}$$

où $\nu=1$ dans la partie du domaine $\Delta$ qui appartient à $G$, et $\nu=-1$ dans le reste du domaine $\Delta$. En tenant compte de l'identité

$$2R^{-1}-r\frac{\partial R^{-1}}{\partial r}=3R^{-1}+r'\frac{\partial R^{-1}}{\partial r'} \tag{4}$$

nous pouvons représenter la fonction $W$ sur la surface $\Sigma$ dans la forme

$$W=\iiint_{\Delta}\nu\left(3R^{-1}+r'\frac{\partial R^{-1}}{\partial r'}\right)d\tau. \tag{5}$$

En substituant (5) dans (1), nous trouvons:

$$I=-\frac{1}{4\pi}\iiint_{\Delta}\nu\left[3\iint_{\Sigma}\left(U\frac{\partial R^{-1}}{\partial n}-R^{-1}\frac{\partial U}{\partial n}\right)d\sigma+r'\frac{\partial}{\partial r'}\iint_{\Sigma}\left(U\frac{\partial R^{-1}}{\partial n}-\right.\right.$$
$$\left.\left.-R^{-1}\frac{\partial U}{\partial n}\right)d\sigma\right]d\tau=\iiint_{\Delta}\nu\left(3U'+r'\frac{\partial U'}{\partial r'}\right)d\tau. \tag{6}$$

Soient ensuite $r=F(\vartheta,\varphi)$ et $r=f(\vartheta,\varphi)$ — les équations des surfaces $S$ et $s$ en coordonnées sphériques, quand l'origine se trouve au point $O$. On peut alors donner à la formule (6) la forme

$$I=\int_0^{2\pi}\int_0^{\pi}\int_f^F\nu\left(3U+r\frac{\partial U}{\partial r}\right)r^2\sin\vartheta\,dr\,d\vartheta\,d\varphi=\int_0^{2\pi}\int_0^{\pi}\nu\left[F^3U(F,\vartheta,\varphi)-\right.$$
$$\left.-f^3U(f,\vartheta,\varphi)\right]\sin\vartheta\,d\vartheta\,d\varphi. \tag{7}$$

En comparant (2) et (7), nous arrivons à la conclusion suivante: pour chaque fonction $U(r,\vartheta,\varphi)$ harmonique dans le domaine fermé $G+G^*$ on a l'égalité

$$\int_0^{2\pi}\int_0^{\pi}\nu\left[F^3U(F,\vartheta,\varphi)-f^3U(f,\vartheta,\varphi)\right]\sin\vartheta\,d\vartheta\,d\varphi=\iint_{\sigma^*}\rho U\,d\sigma. \tag{8}$$

Désignons ensuite par $\overline{U}(r,\vartheta,\varphi)$ une fonction égale à $\nu$ sur la surface $S$ et harmonique à l'intérieur de cette surface. Alors

$$\int_0^{2\pi}\int_0^{\pi}\nu\left[F^3\overline{U}(F,\vartheta,\varphi)-f^3\overline{U}(f,\vartheta,\varphi)\right]\sin\vartheta\,d\vartheta\,d\varphi\geqslant$$
$$\geqslant\int_0^{2\pi}\int_0^{\pi}(F^3-f^3)\sin\vartheta\,d\vartheta\,d\varphi=3T \tag{9}$$

parce que $|\overline{U}|\leqslant 1$ à l'intérieur de $S$. D'ailleurs, dans (9) l'égalité n'a lieu que si les domaines $G$ et $G^*$ coïncident, ou bien si $\nu=\text{const}$ sur $S$ (c'est-à-dire, dans le cas où l'un des domaines $G$ et $G^*$ contient l'autre à son intérieur). Ainsi il existe toujours une fonction $U$ vérifiant les conditions

$$|U|\leqslant 1 \quad \text{sur } \sigma^*, \quad \iint_{\sigma^*}\rho U\,d\sigma\geqslant 3T \tag{10}$$

d'où

$$T\leqslant\frac{1}{3}\iint_{\sigma^*}|\rho|\,d\sigma. \tag{11}$$

Prenons maintenant une surface fermée et unie quelconque $\Sigma^*$, qui contient entièrement à son intérieur la surface $\sigma^*$, et une fonction $\delta$ arbitraire bornée sur $\Sigma^*$. Nous trouvons

$$\iint_{\Sigma^*} \delta \frac{\partial W}{\partial n} d\sigma = \iint_{\sigma^*} \rho U^* d\sigma \tag{12}$$

où

$$U^* = \iint_{\Sigma^*} \delta \frac{\partial R^{-1}}{\partial n} d\sigma.$$

En approchant la surface $\Sigma^*$ à la surface $\sigma^*$, nous arrivons à la conclusion suivante. Pour chaque fonction $\delta$ bornée sur $\sigma^*$ on a l'égalité

$$\iint_{\sigma^*} \delta \frac{\partial W}{\partial n} d\sigma = \iint_{\sigma^*} \rho U^* d\sigma \tag{13}$$

où $U^*$ est le potentiel intérieur de double couche de densité $\delta$ sur la surface $\sigma^*$.

Désignons maintenant par $\bar{\delta}$ la densité de la double couche pour laquelle

$$U^* = \eta \qquad \text{sur } \sigma^* \tag{14}$$

où $\eta = 1$ sur la partie de la surface $\sigma^*$, où la densité $\rho$ est positive et $\eta = -1$ sur la partie restante de cette surface. La fonction $\bar{\delta}$ est déterminée par l'équation intégrale du problème de Dirichlet

$$\bar{\delta}(P) + \frac{1}{2\pi} \iint_{\sigma^*} \bar{\delta}(P')\, d\omega(P, P') = \frac{1}{2\pi} \eta(P) \tag{15}$$

où $d\omega$ est l'angle solide sous lequel on voit l'élément de la surface $\sigma^*$ contenant le point $P'$ du point $P$ sur la surface, donc $\bar{\delta}$ aura la forme

$$\bar{\delta} = \frac{\eta}{2\pi} + \text{une fonction continue sur } \sigma^*. \tag{16}$$

D'autre part, il suit de (11) et (13)

$$T \leqslant \frac{1}{3} \iint_{\sigma^*} \bar{\delta} \frac{\partial W}{\partial n} d\sigma \tag{17}$$

ce qui démontre le théorème énoncé.

Remarque. En tenant compte de l'inégalité (11), on peut obtenir une évaluation de la proximité des domaines $G$ et $G^*$ à partir des valeurs de la fonction $V$ et de ses dérivées premières sur un certain plan $D$. situé hors de $\sigma^*$, en résolvant l'équation régulière linéaire intégrale

$$\iint_{\sigma^*} \frac{\rho d\sigma}{R} = W \qquad \text{sur } D. \tag{18}$$

Si l'équation (18) possède une solution, on peut en conclure que le domaine $\sigma^*$ contient tous les points singuliers de la fonction $V$, et obtenir l'évaluation cherchée d'après la formule (11).

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*MATHÉMATIQUES*

# APPLICATION DE LA TRANSFORMATION DE LAPLACE AUX PROBLÈMES LIMITES POUR L'ÉQUATION DES ONDES

Par S. MICHLIN

*(Présenté par N. E. Kotchin, de l'Académie, le 10. II. 1941)*

§ 1. Etant donnée l'équation des ondes

$$\frac{\partial^2 U}{\partial x^2}+\frac{\partial^2 U}{\partial y^2}-\frac{\partial^2 U}{\partial t^2}=0 \tag{1}$$

posons le problème suivant: trouver une solution $U$ définie pour $t \geqslant 0$ dans un domaine $B$ du plan $XOY$, borné par un contour fermé $L$, ou, ce qui est le même, à l'intérieur d'un demi-cylindre $G$ dont les génératrices sont parallèles à l'axe $OT$ et la directrice est $L$. Le contour $L$ est supposé ayant une courbure continue. La fonction $U$ cherchée vérifie les conditions:

1) Dans le domaine

$$U(x, y, 0)=F(x, y); \quad U_t(x, y, 0)=\varphi(x, y). \tag{2}$$

2) Sur le demi-cylindre $G$ (pour $t \geqslant 0$)

$$U=\psi(x, y, t) \tag{3}$$

où $F$, $\varphi$, $\psi$ sont des fonctions connues *.

Désignons par $\sigma'$ la partie de la surface du demi-cylindre $G$, coupée par le cône caractéristique

$$t_0-t=r; \quad r^2=(x_0-x)^2+(y_0-y)^2. \tag{4}$$

Si le point $M_0(x_0, y_0, t_0)$ est situé sur $G$, alors $\sigma'$ sera désigné par $\sigma_0$. Nous allons chercher la solution de notre problème dans la forme [1]:

$$U(M_0)=A(M_0)+\frac{1}{2\pi}\int_{\sigma'}\int Q_t(M)\frac{(t_0-t)\cos(n, r)}{r\sqrt{(t_0-t)^2-r^2}}dS \tag{5}$$

où $A$ est une fonction connue, $Q$ — la fonction inconnue nulle pour $t=0$; $n$ est la normale extérieure à $L$. En faisant le point $M_0$ tendre vers $G$, nous obtenons une équation intégro-différentielle

$$Q(M_0)-\frac{1}{\pi}\int_{\sigma_0}\int Q_t(M)\frac{\cos(n, r)}{r}\frac{(t_0-t)\,dS}{\sqrt{(t_0-t)^2-r^2}}=$$
$$=2\psi(M_0)-2A(M_0)=A(M_0). \tag{6}$$

* La condition (3) peut être remplacée par la suivante: $\frac{dU}{dn}=\chi(x, y, t)$. Cela n'aura pas d'influence sur la méthode. Cette remarque subsiste pour le § 2.

Désignons par $s$ la longueur de l'arc du contour $L$. Posons encore

$$\int_0^\infty Q(M)\, e^{-\tau t}\, dt = \Phi(s, \tau), \quad \int_0^\infty A(M)\, e^{-\tau t}\, dt = B(s, \tau). \tag{7}$$

En appliquant à (16) la transformation de Laplace, nous obtenons pour la fonction inconnue $\Phi(s, \tau)$ une équation du type de Fredholm:

$$\Phi(s_0, \tau) - \frac{1}{\pi} \int_L \frac{\cos(n, r)}{r} K(r\tau)\, \Phi(s, \tau)\, ds = B(s_0, \tau). \tag{8}$$

Ici

$$K(x) = x \int_1^\infty \frac{v e^{-vx}}{\sqrt{v^2 - 1}}\, dv. \tag{9}$$

Pour les valeurs de $x$ voisines de zéro on aura

$$K(x) = 1 + \frac{1}{2} x^2 \ln x + \ldots \tag{10}$$

et le noyau de l'équation (8) est continu.

Si $R(\tau) \to +\infty$, on a

$$\int_L \int_L \left| \frac{\cos(n, r)}{r} K(r\tau) \right|^2 ds_0\, ds \to 0.$$

Il en résulte que l'équation (8) est résoluble par la méthode des approximations successives, si la partie réelle de $\tau$ est assez grande. Il existe une droite $R(\tau) = x$, le long de laquelle le déterminant de Fredholm de l'équation (8) est différent de zéro. En résolvant l'équation (8) et prenant l'inversion d'après la formule

$$Q(M) = \frac{1}{2\pi i} \int_{k-i\infty}^{k+i\infty} \Phi(s, \tau)\, e^{\tau t}\, d\tau \tag{11}$$

nous aurons en même temps une solution de notre problème limite.

§ 2. On peut poser un problème analogue à celui du § 1 pour l'équation

$$\frac{\partial^2 U}{\partial x^2} + \frac{\partial^2 U}{\partial y^2} + \frac{\partial^2 U}{\partial z^2} - \frac{\partial^2 U}{\partial t^2} = 0. \tag{12}$$

Dans ce cas il faut trouver une solution de l'équation (12) dans un domaine $D$ de l'espace $(x, y, z)$ borné par une certaine surface fermée $S$*. Cette solution vérifie les conditions de Cauchy pour $t = 0$ et prend des valeurs données sur $S$.

Considérons la sphère de rayon $t_0$ et de centre $M_0(x_0, y_0, z_0)$. Désignons par $S'$ la partie de cette sphère, coupée par la surface $S$ et située à l'intérieur de $D$, et par $\sigma'$ — la partie de la surface $S$, coupée par la sphère indiquée. En appliquant la méthode connue de Kirchhoff [voir, par exemple ([2])], nous aurons

$$U(M_0, t_0) = \frac{1}{4\pi} \int_{S'+\sigma'} \int \left\{ \left[ U(M, t_0) \right] \frac{\cos(n, r)}{r^2} - \frac{1}{r} \left[ \frac{\partial U(M, t_0)}{\partial n} \right] - \right.$$

$$\left. - \frac{\cos(n, r)}{r} \left[ \frac{\partial U(M, t_0)}{\partial t_0} \right] \right\} dS. \tag{13}$$

Ici $n$ est la direction de la normale extérieure à $S$ au point $M$; les crochets [ ] désignent la valeur retardée de la fonction, de sorte que pour une fonction $W$ quelconque on a

$$[W(M, t_0)] = W(M, t_0 - r).$$

* Nous supposerons que cette surface vérifie les conditions de Liapounoff.

La formule (13) nous donne la solution du problème, si l'on connaît la fonction

$$P(M, t) = \frac{\partial U(M, t)}{\partial n} \tag{14}$$

sur la surface $S$.

Transportons le point $M_0$ sur la surface $S$ et dérivons (13) le long de la normale $n_0$ à la surface $S$ au point $M_0$. En tenant compte des données du problème, nous aurons pour la quantité $P(M, t)$ une équation intégro-différentielle

$$P(M_0, t_0) + \frac{1}{2\pi} \int\int_{\sigma_0} \frac{\cos(n_0, r)}{r^2} \int_0^{t_0 - r} P_t(M, t)\, dt\, dS -$$

$$- \frac{1}{2\pi} \int\int_{\sigma_0} \frac{\cos(n_0, r)}{r} \int_0^{t_0 - r} P_{tt}(M, t)\, dt\, dS = C(M_0, t_0). \tag{15}$$

Ici $C$ est une fonction connue; par $\sigma_0$ on désigne la surface $\sigma'$ quand le point $M_0$ se trouve sur la surface $S$. Posons

$$\int_0^\infty e^{-\tau t} P(M, t)\, dt = \Psi(M, \tau). \tag{16}$$

En appliquant à (15) la transformation de Laplace, nous obtenons une équation du type de Fredholm

$$\Psi(M_0, \tau) + \frac{1}{2\pi} \int\int_S \frac{\cos(n_0, r)}{r^2} e^{-\tau r} (1 + \tau r)\, \Psi(M, \tau)\, dS = E(M_0, \tau). \tag{17}$$

En appliquant l'itération à l'équation (17), nous aurons une nouvelle équation dont le noyau est à carré sommable. Il est facile à vérifier que pour $R(t) \longrightarrow +\infty$ l'intégrale du carré du module du noyau itéré tend vers zéro. Il en résulte que si $R(\tau)$ est assez grand, l'équation (17) est résoluble par la méthode des approximations successives. On peut appliquer les raisonnements du § 1.

L'extension de cette méthode au cas d'un nombre quelconque de variables ne présente aucune difficulté.

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*HYDROTECHNIK*

# ÜBER DIE BEWEGUNGSSTABILITÄT DES ELASTISCH AUFGEHÄNGTEN SENKRECHTEN SCHIEBERS DER SCHLEUSENGALLERIEN

Von I. A. TSCHARNY

*(Vorgelegt von N. E. Kotschin, Mitglied der Akademie, den 14. II. 1941)*

1. Es ist bekannt, dass mitunter eine intensive Schwingung der Schieberverschlussanordnung spontan entstehen kann, die von schädlicher Einwirkung auf den Bau ist. Die Schwingung der Verschlusschieber in der offenen Gerinnen ist von Müller (1) untersucht worden, der festgestellt hat, dass die Schwingungen durch das periodisch auftretende Abreissen und Festkleben der Grenzschicht an der unteren Kante des Schiebers hervorgerufen werden, die bei einer nicht sachgemässen Ausführung des Umrisses der unteren Kante des Schieberverschlusses stattfinden können.

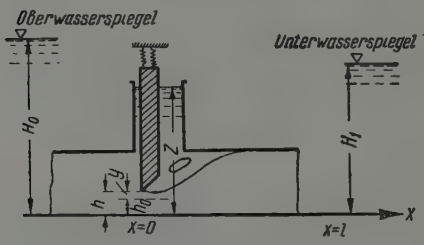

Abb. 1.

Dagegen sind, soweit uns bekannt, die Schwingungen des Schieberverschlusses in den Schleusengallerien noch nicht studiert worden. Im hydrodynamischen Laboratorium des Energetischen Instituts ist diese Frage aufgenommen worden, und ihre experimentelle und theoretische Erforschung wird gegenwärtig fortgesetzt. Einige theoretische Ergebnisse werden hiermit mitgeteilt.

2. Das Schema der Anordnung des Verschlusschiebers in der Schleusengallerie ist in der Abb. 1 widergegeben. Um das periodische Abreissen und Festkleben der Grenzschicht zu verhindern, ist der unteren Kante des Schiebers die Form eines Keiles gegeben worden. Der Schieber ist in einen Schacht eingebaut, dem die Aufgabe des Dämpfers der Druckpulsation zugeteilt ist.

3. Als Ausgangsdifferentialgleichungen für die Aufstellung des Geschwindigkeits- und der Druckverteilung in der Gallerie, dienen die bekannten Gleichungen des hydraulischen Stosses:

$$-\frac{\partial p}{\partial x}=\rho_0\frac{\partial v}{\partial t}, \quad -\frac{\partial p}{\partial t}=K\frac{\partial v}{\partial x} \tag{1}$$

wo $v(x,t)$, $p(x,t)$ — die mittleren Geschwindigkeiten und Drücke im Querschnitt, $\rho$ — die Dichte, $K$ — der Modul der Elastizität für Wasser ist, der zusammenfassend die Wirkung der Kompressibilität der Flüssigkeit und die Elastizität der Galleriewände berücksichtigt. Hier ist von der Berücksichtigung des hydraulischen Widerstandes und der Ungleichmässigkeit der Geschwindigkeitsverteilung im Querschnitt Abstand genommen. Eine tiefer gehende Untersuchung (²) bestätigt die Zulässigkeit einer solchen Vernachlässigung für die gebräuchlichen Abmessungen und Längen der Schleusengallerien.

4. Die Grenzbedingungen erhält man folgenderweise: $x=0$ (beim Schieber):

$$Q - fv = F\frac{dz}{dt} \tag{2}$$

wo $Q$ — die in der Zeiteinheit unter dem Schieber abfliessende Wassermenge, $f$ — die Querschnittsfläche der Gallerie, $F$ — die Querschnittsfläche des Schachtes, $z$ — die Höhe des Wasserspiegels in dem Schachte sind.

$x=l$ (beim Auslauf in den Unterwasserspiegel): $p=\text{const.}$ (3)

Weiterhin erhält man bei $x=0$ (Abb. 1):

$$H_0 - z = \xi\frac{Q^2}{2gb^2h^2} \tag{4}$$

wo $\xi$ — der Widerstandskoeffizient, $b$ — die Breite der Gallerie (rechtwinklich), $h$ — die Schieberöffnung im gegebenen Zeitmomente sind.

Sei $Q=Q_0+q$; $v=v_0+v'$; $z=z_0+z'$; $h=h_0+y$, wo $Q_0$, $v_0$, $z_0$, $h_0$ die stationären Werte, $q$, $v'$, $z'$, $y$ ihre kleinen Zuwächse sind. Dann ist

$$q - fv' = F\frac{dz'}{dt}\,. \tag{2,1}$$

Indem wir die zweiten Potenzen der Zuwächse vernachlässigen, erhalten wir aus (4):

$$-z' = 2(H_0 - z_0)\left(\frac{q}{Q_0} - \frac{y}{h_0}\right). \tag{4,1}$$

Zu diesen Gleichungen muss noch die Gleichung der Bewegung des Schiebers als festen Körpers zugefügt werden:

$$M\ddot{y} + \eta\dot{y} + ay = sp' \tag{5}$$

wo $M$ — die Masse des Schiebers; $a$ — die Federzahl; $\eta\dot{y}$ — die Reibungskraft, die im Verhältnis zur ersten Potenz der Geschwindigkeit angenommen wird; $s$ — die Querschnittsfläche des Schiebers ($s=bd$, $d$ — die Dicke des Schiebers); $p'$ — die den stationären Druck überhöhende Druckvergrösserung. Augenscheinlich ist

$$p' = \gamma z' \tag{6}$$

wo $\gamma$ das spezifische Gewicht des Wassers ist.

Aus den Gleichungen (2, 1), (4, 1), (5) und (6) erhalten wir nach einfachen Berechnungen die Grenzbedingung für $x=0$

$$h_0\left(\frac{f\ddot{v}}{Q_0} + \frac{F}{\gamma Q_0}\ddot{p} + \frac{\ddot{p}}{2p_0}\right) + 2\nu h_0\left(\frac{f\dddot{v}}{Q_0} + \frac{F}{\gamma Q_0}\dddot{p} + \frac{\dddot{p}}{2p_0}\right) + \\ + \omega_0^2 h_0\left(\frac{fv}{Q_0} + \frac{F}{\gamma Q_0}\dot{p} + \frac{\dot{p}}{2p_0}\right) = \frac{gsp}{G} \tag{7}$$

wo

$$2\nu = \frac{\eta}{M}; \quad \omega_0^2 = \frac{a}{M}; \quad p_0 = \gamma (H_0 - z_0) \tag{8}$$

$G$ das Gewicht des Schiebers und $g = 9.81$ m/sec$^2$ sind. Wie gewöhnlich, bedeuten die Punkte über den Buchstaben die Differentiation nach der Zeit, und unter $p$ und $v$ werden hier, wie auch weiterhin, nur die veränderlichen Komponenten des Druckes und der Geschwindigkeit verstanden.

5. Die Definition der Bewegungsstabilität werden wir folgendermassen durchführen: wir bestimmen das Spektrum der Eigenfrequenzen des Systems Strömung — Schieber, indem wir uns diese Frequenzen als komplexe Grössen denken. Haben die Frequenzen einen positiven imaginären Teil, so wird die Bewegung stabil sein; sind aber die Frequenzen mit negativem imaginärem Teil, so wird die Bewegung unstabil. Zu diesem Zweck wenden wir uns an die Gleichungen (1). Partikuläre Lösungen für $v$ und $p$ können in der Form:

$$\left.\begin{aligned} v &= \left(A_1 \cos\frac{\omega x}{c} + A_2 \sin\frac{\omega x}{c}\right) e^{i\omega t} \\ p &= i\rho c \left(A_2 \cos\frac{\omega x}{c} - A_1 \sin\frac{\omega x}{c}\right) e^{i\omega t} \end{aligned}\right\} \tag{9}$$

gesucht werden, wo $c = \sqrt{\frac{K}{\rho}}$ — die Schallgeschwindigkeit in der Flüssigkeit, $A_1$, $A_2$ — Funktionen der komplexen Frequenz $\omega$ sind. Aus den Grenzbedingungen (3) und (7) erhalten wir:

$$\rho c\left[\frac{F}{\gamma Q_0}\omega^3 - i\left(\frac{1}{2p_0} + \frac{2\nu F}{\gamma Q_0}\right)\omega^2 - \left(\frac{\nu}{p_0} + \frac{\omega_0^2 F}{\gamma Q_0}\right)\omega + i\left(\frac{\omega_0^2}{2p_0} - \frac{gs}{Gh_0}\right)\right] A_2 -$$
$$- \frac{f}{F}(\omega^2 - i2\nu\omega - \omega_0^2) A_1 = 0; \quad A_2 \cos\frac{\omega l}{c} - A_1 \sin\frac{\omega l}{c} = 0. \tag{10}$$

Wird die Determinante aus den Koeffizienten der Grössen $A_2$ und $A_1$ des Systems (10) gleich 0 gesetzt, so erhalten wir, nach der Überleitung zur dimensionslosen Form, folgende Gleichung für die Eigenfrequenz des Systems:

$$\Phi(\varphi) = (a_3\varphi^3 - ia_2\varphi^2 - a_1\varphi + ia_0)\sin\varphi - (\varphi^2 - ib_1\varphi - b_0)\cos\varphi = 0 \tag{11}$$

wo

$$\left.\begin{aligned} &\varphi = \frac{\omega l}{c}; \quad a_3 = \frac{F}{f}\frac{c^2}{gl}; \quad a_2 = \frac{c\nu v_0}{2p_0} + \frac{2F}{f}\frac{\nu c}{g}; \quad a_1 = \frac{\gamma l}{p_0}\frac{\nu v_0}{g} + \frac{F}{f}\frac{l\omega_0^2}{g}; \\ &a_0 = \frac{c\nu v_0}{2p_0}\varphi_0^2 - \frac{\gamma s l}{G}\frac{l}{h_0}\frac{v_0}{c}; \quad b_1 = 2\frac{\nu l}{c}; \quad b_0 = \varphi_0^2; \quad \varphi_0 = \frac{\omega_0 l}{c} \end{aligned}\right\} \tag{12}$$

wo $v_0 = \frac{Q_0}{f}$ die mittlere Geschwindigkeit der Strömung in der Gallerie ist.

In allen den Fällen, wo die Wurzeln der Gleichung (11) in der oberen Halbebene zu liegen kommen, ist die Bewegung stabil. Fallen aber einige Wurzeln auf die untere Halbebene, so ist die Bewegung, allgemein gesagt, nicht stabil, obgleich zu erwarten ist, dass eine praktische Bedeutung nur die Wurzeln mit dem kleinsten reellen Teil haben werden, d. h. die den Frequenzen entsprechen, die mit $\omega_0$ der Eigenfrequenz der Schwingungen des Schiebers vergleichbar sind.

Die analytische Erforschung der Wurzelverteilung der der (11) ähnlichen Gleichungen führt zur Betrachtigung einer unendlichen Reihenfolge der Determinanten von Gurwitz (3). Scheinbar werden deswegen hier reine Errechnungsmethoden sich als praktisch brauchbar erweisen, die für die gegebenen Werte der Parameter $a_0$, $a_1$, $a_2$, $a_3$, $b_0$, $b_1$ die Stabilität oder die Unstabilität festzustellen gestatten. Als eine dieser Methoden

kann folgendes empfohlen werden: wir führen eine neue Funktion $\zeta$ ein, die mit $\varphi$ durch das Verhältnis $\varphi = i\frac{\zeta+1}{\zeta-1}$ in Verbindung steht. Der unteren Halbebene $\varphi$ entspricht auf der Ebene $\zeta$ eine Kreiskurve des Einheitshalbmessers mit dem Zentrum in dem Koordinatenanfang. Dann erhält man anstatt der Gleichung (11) folgende Gleichung

$$\Phi\left(i\frac{\zeta+1}{\zeta-1}\right) = W = u + iv. \qquad (13)$$

Mit Hilfe der Gleichung (13) ist es möglich auf der Ebene $W$ ein Kontur zu entwerfen, der der Kreiskurve des Einheitshalbmessers auf der Ebene $\zeta$ oder, was dasselbe ist, der unteren Halbebene von $\varphi$ entspricht. Nach der Lage des Punktes $W=0$, in Beziehung zu dem entworfenen Kontur, kann man über das Nichtvorhanden- oder Vorhandensein der Wurzeln auf der unteren Halbebene von $\varphi$ und damit über die Stabilität oder die Unstabilität der Bewegung urteilen. Diese Methode ist für beliebige Werte der Parameter $a_0$, $a_1$, $a_2$, $a_3$, $b_0$, $b_1$ verwertbar.

Praktische Notwendigkeit in einer solchen Berechnung entsteht aber nur dann, wenn die Parameter $a_0$, $a_1$, $a_2$, $a_3$ und $b_0$, $b_1$ sich als Werte derselben Grössenordnung erweisen. In Wirklichkeit sind die Parameter $a_3$ und $a_2$ für die vorhandenen Gefälle und Abmessungen der Schleusengallerie viel grösser als alle anderen. Daraus folgt, dass die Wurzeln der Gleichung (11) sich in der Nähe der Punkte $\varphi = n\pi$, $n=0, 1, 2,$ befinden. In diesem Falle kann gesetzt werden

$$\varphi \cong n\pi + \delta \qquad (14)$$

wo $n=0, 1, 2$ und $\delta$ eine sehr kleine reelle oder komplexe Zahl ist. Dann erhält man angenähert aus den Gleichungen (12), (10) und (9):

$$\omega = \frac{(n\pi+\delta)c}{l} \quad \text{und} \quad v_{x=0} = A_1 e^{\frac{i(n\pi+\delta)ct}{l}} \qquad (14,1)$$

$$A_2 = A_1 \operatorname{tg}\varphi = A_1 \operatorname{tg}\delta \cong A\delta,$$

woraus

$$p_{x=0} = i\rho c A_1 \delta e^{i(n\pi+\delta)\frac{ct}{l}} = i\rho c A_1 (n\pi+\delta) e^{i(n\pi+\delta)\frac{ct}{l}} - i\rho c n\pi A_1 e^{i(n\pi+\delta)\frac{ct}{l}}. \qquad (15)$$

Wenn wir die Gleichung (14,1) berücksichtigen, so können wir Gleichung (15) in der Form darstellen:

$$p_{x=0} = \rho l \left(\frac{\partial v}{\partial t}\right)_{x=0} - in\pi\rho c\,(v)_{x=0}. \qquad (16)$$

Für die niedrigste Frequenz ($n=0$) erhalten wir die bekannte Gleichung für die unkompressible Flüssigkeit. Setzen wir nun die Gleichung (16) in die Grenzbedingungen (7) ein, so erhalten wir, nach der Trennung der reellen und imaginären Teile, zwei gewöhnliche Differenzialgleichungen der vierten und der dritten Ordnung mit konstanten Koeffizienten, deren stabile Lösungen leicht mittels Determinanten von Gurwitz gefunden werden können.

Für die kleinste Frequenz ($n=0$) erhalten wir nach Vereinfachungen aus (7):

$$\frac{F}{f}\frac{h_0 l}{g v_0}\dddot{v} + \frac{h_0 l}{g}\left[\frac{1}{2(H_0-z_0)} + 2\frac{\nu}{v_0}\frac{F}{f}\right]\ddot{v} + h_0\left[\frac{l}{g}\left(\frac{\nu}{(H_0-z_0)} + \frac{\omega_0^2 F}{v_0 f}\right) + \frac{1}{v_0}\right]\dot{v} + \\ + \left[2\nu\frac{h_0}{v_0} + \frac{h_0\omega_0^2 l}{2g(H_0-z_0)} - \frac{\gamma s l}{G}\right]\dot{v} + \frac{h_0\omega_0^2}{v_0}v = 0. \qquad (17)$$

Bei Auflösung der Determinanten von Gurwitz, erhalten wir folgende Kriteriums der Stabilität der durch die Gleichung (17) definierten Bewegung:

$$2\frac{\nu h_0}{v_0}+\frac{h_0\omega_0^2 l}{2g(H_0-z_0)}-\frac{\gamma s l}{G}>0 \tag{18}$$

$$1>-\frac{\frac{F}{f}\left(\frac{2\nu h_0}{v_0}+\frac{h_0\omega_0^2}{2g}\frac{l}{H_0-z_0}-\frac{\gamma s l}{G}\right)}{\left[\frac{h_0}{2(H_0-z_0)}+\frac{F}{f}\frac{2\nu h_0}{v_0}\right]\cdot\left[\frac{l}{H_0-z_0}\frac{\nu v_0}{g}+\frac{F}{f}\frac{l\omega_0^2}{g}+1\right]}+$$

$$+\frac{\frac{h_0\omega_0^2}{g}\left[\frac{l}{2(H_0-z_0)}+\frac{F}{f}\frac{2\nu l}{v_0}\right]}{\left(\frac{l}{H_0-z_0}\frac{\nu v_0}{g}+\frac{F}{f}\frac{l\omega_0^2}{g}+1\right)\left(\frac{2\nu h_0}{v_0}+\frac{h_0\omega_0^2}{2g}\frac{l}{H_0-z_0}-\frac{\gamma s l}{G}\right)}. \tag{19}$$

Die Untersuchung der Gleichungen (18) und (19) für verschiedene Bedingungen ist in (2) angeführt. Es kann nachgewiesen werden, dass auch für $n\neq 0$, also für höhere Frequenzen, die durch die Wirkung der Kompressibilität der Flüssigkeit bedingt sind, keine zusätzlichen Kriteriums für die Stabilität gewonnen werden. Folglich, ist bei Nichtbeachtung der Gleichungen (18) und (19) die Möglichkeit für unerwünschte Eigenschwingungsbewegungen des Schiebers vorhanden.

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*Übersetzt von K. Lehmann.*

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*THÉORIE DE L'ÉLASTICITÉ*

# SUR LES TENSIONS DANS UNE PLAQUE ELLIPTIQUE

Par D. SCHERMAN

*(Présenté par S. L. Soboleff, de l'Académie, le 10. II. 1941)*

Le problème de la détermination des tensions dans une plaque elliptique à partir des forces extérieures sur la frontière a été résolu par N. L. Mouskhelichvili (1). Nous donnons dans ce qui suit une autre solution du même problème, qui permet d'obtenir des formules plus simples pour le calcul. D'ailleurs, il nous semble que la méthode proposée peut aussi être utilisée dans quelques autres problèmes de la théorie de l'élasticité.

Nous partons des équations intégrales de Lauricella (2), qu'il a obtenues en résolvant le problème sur les tensions dans un domaine simplement connexe. Ces équations écrites sous la forme complexe, sont (3)

$$\omega(t_0)+\frac{1}{2\pi i}\int_L \omega(t)\,d\,\lg\frac{t-t_0}{\bar{t}-\bar{t}_0}-\frac{1}{2\pi i}\int_L \overline{\omega(t)}\,d\,\frac{t-t_0}{\bar{t}-\bar{t}_0}=f(t_0) \qquad (1)$$

où la courbe $L$ est une ellipse bornant un domaine situé dans le plan $z=x+iy$; $t$ et $t_0$ sont les affixes des points de $L$; la fonction $f(t_0)=$ $=i\int_0^{s_0}(X_n+iY_n)\,ds$, où $X_n+iY_n$ est le vecteur des tensions extérieures et $s$ est l'arc de $L$, compté à partir d'une certaine origine fixe.

Après avoir déterminé $\omega(t)$, on trouve les fonctions de la variable complexe $\varphi(x)$ et $\psi(z)$, introduites par G. V. Kolossoff au moyen des formules (1):

$$\left.\begin{aligned}\varphi(z)&=\frac{1}{2\pi i}\int_L \frac{\omega(t)}{t-z}\,dt,\\ \psi(z)&=\frac{1}{2\pi i}\int_L \frac{\overline{\omega(t)}}{t-z}\,dt-\frac{1}{2\pi i}\int_L \frac{\bar{t}\,\omega'(t)}{t-z}\,dt.\end{aligned}\right\} \qquad (2)$$

Faisons la représentation de l'ellipse $L$ sur une circonférence $\gamma$ de rayon $\rho>1$ dans le plan $\zeta$ au moyen de la formule $t=R\left(\sigma+\frac{1}{\sigma}\right)$, où $\sigma$ est l'affixe de $\gamma$.

Alors, en posant $\omega(t)=\omega^*(\sigma)$, $f(t)=f^*(\sigma)$, on transforme l'égalité (2) dans la forme

$$\omega^*(\sigma_0)+\frac{1}{2\pi i}\int_\gamma \omega^*(\sigma)\,\frac{d\sigma}{\sigma-\frac{1}{\sigma_0}}-\frac{1}{2\pi i}\int_\gamma \omega^*(\sigma)\,\frac{d\sigma}{\sigma-\frac{\rho^4}{\sigma_0}}-$$

$$-(\rho^4-1)\frac{1}{2\pi i}\int_\gamma \frac{\overline{\omega^{*\prime}(\sigma)}}{\sigma}\,\frac{d\sigma}{\sigma-\frac{\rho^4}{\sigma_0}}=f^*(\sigma_0). \qquad (3)$$

On cherche la fonction inconnue $\omega^*(\sigma)$ sous la forme d'une série de Fourier $\omega^*(\sigma)=\sum_{-\infty}^{\infty} a_k\sigma^k$. Alors, en posant $f^*(\sigma)=\sum_{-\infty}^{\infty} A_k\sigma^k$ et en remarquant que $\left|\frac{1}{\sigma_0}\right|<1<\rho$, $\left|\frac{\rho^4}{\sigma_0}\right|>\rho$, on obtient de (3):

$$2a_0+\sum_1^\infty\left[a_k+a_{-k}\rho^{-4k}+(\rho^4-1)\,k\overline{a}_k\rho^{-2(k+1)}\right]\sigma^k+\sum_{-1}^{-\infty}(a_k+a_{-k})\sigma^k=$$

$$=\sum_{-\infty}^{\infty}A_k\sigma^k. \qquad (4)$$

D'où, en égalant les coefficients des mêmes puissances de $\sigma$, nous aurons:

$$\left.\begin{aligned} 2a_0=A_0,\quad a_k+a_{-k}\rho^{-4k}+(\rho^4-1)\,k\overline{a}_k\rho^{-2(k+1)}=A_k,\\ a_k+a_{-k}=A_{-k}.\quad (k=1,2,\ldots,\infty)\end{aligned}\right\} \qquad (5)$$

La seconde de ces dernières égalités, quand on exclut $a_{-k}$ prend la forme

$$(1-\rho^{-4k})\,a_k+\rho^{-2(k+1)}(\rho^4-1)\,k\overline{a}_k=A_k-\rho^{-4k}A_{-k}.\quad (k=1,2,\ldots,\infty) \qquad (6)$$

En posant dans cette égalité $k=1$, on trouve

$$a_1+\overline{a}_1=\frac{A_1-\rho^{-4}A_{-1}}{1-\rho^{-4}}. \qquad (7)$$

Le second membre de cette dernière égalité est une quantité réelle. Ainsi qu'il est facile à vérifier, cela résulte de ce que le moment principal des forces extérieures s'annulle. La partie imaginaire $a_1$ peut être considérée comme nulle. On trouve de l'équation (6) tous les $a_k$, et ensuite de la troisième des équations (5) tous les $a_{-k}$.

Maintenant en s'adressant, par exemple, à la première des équations (2), on trouve sans peine la valeur de la fonction $\varphi(t)=\varphi^*(\sigma)$ sur la circonférence $\gamma$:

$$\varphi^*(\sigma)=a_0+\sum_1^\infty a_k\left(\sigma^k+\frac{1}{\sigma^k}\right). \qquad (8)$$

L'expression pour la fonction $\psi(t)=\psi^*(\sigma)$ peut être trouvée d'une manière analogue à partir de la seconde des équations (2), ou bien à partir des conditions sur la frontière $L$.

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*THÉORIE DE L'ÉLASTICITÉ*

# SUR LA GÉNÉRALISATION POSSIBLE DE LA MÉTHODE ITÉRATIVE DE DESCRIPTION DES PHÉNOMÈNES DES ACTIONS ULTÉRIEURES ET DE L'HYSTÉRESIS, ET DE LA CONCORDANCE DE CETTE MÉTHODE AVEC CELLE DE BOLTZMAN-VOLTERRA

Par M. V. MATCHINSKY
*(Présenté par N. N. Lusin, de l'Académie, le 24. I. 1941)*

§ 1. En 1938 il a été proposé par V. I. Kovalenkoff et par l'auteur la méthode de description de tout phénomène des actions ultérieures en général, et de l'hystéresis en particulier (1). Cette méthode mathématique se réduisait essentiellement au suivant. Soit étudiée la dépendance de deux grandeurs physiques $y$ et $x$ (le champ magnétique $B$ et l'induction $H$, la charge $X$ et la déformation $x_{ij}$, etc. *).

Supposons que la courbe primaire soit décrite par la fonction

$$y = f(x) \tag{1}$$

et que l'on change, pour $x$ égal à

$$x_{1,n},\ x_{2,n},\ x_{3,n},\ x_{4,n},\ \text{etc.} \tag{2}$$

le sens du processus (c'est-à-dire, on augmente d'abord $x$ de 0 à $x_{1,n}$ et puis on le diminue de $x_{1,n}$ jusqu'à $x_{2,n}$, après quoi, de nouveau, on l'augmente de $x_{2,n}$ à $x_{3,n}$, etc. Désignons par

$$\begin{gathered}\varphi_0(x) = x;\ \varphi_1(x) = \varphi(x, x_{1,n});\\ \varphi_2(x) = \varphi[\varphi(x, x_{1,n}), x_{2,n}];\ \varphi_3(x) = \varphi\{\varphi[\varphi(x, x_{1,n}), x_{2,n}], x_{3,n}\}.\end{gathered} \tag{3}$$

Alors, la forme de la relation entre $y$ et $x$ s'exprimera

$$\left.\begin{aligned} y &= f(x) && \text{pour } 0 < x < x_{1,n};\\ y &= \varphi_1(f(x)) && \text{pour } x_{2,n} < x < x_{1,n};\\ y &= \varphi_2(f(x)) && \text{pour } x_{3,n} < x < x_{3,n},\end{aligned}\right\} \tag{4}$$

etc. jusqu'à un renversement quelconque du processus, et l'itération.

La méthode en question a été proposée dans les articles de V. I. Kovalenkoff et de l'auteur cités, principalement pour raisons de simplicité (par rapport au procédé de Boltzman-Volterra). D'autre part, l'étude minutieuse des courbes expérimentales montre que dans toute une série de cas on a besoin d'une généralité plus grande, que celle des équations (4).

Une généralisation utile de la dernière équation (4) pourrait être obtenue de la manière suivante: supposons que la courbe primaire

* D'autres exemples sont donnés dans notre article cité.

soit décrite, comme auparavant, par la fonction (1). Soit déduite par les fonctions analogues à (3), au surplus de la fonction $\varphi$, une autre fonction $\psi(x)$, de même que ses itérations $\psi_2$, $\psi_3$, etc. Alors le caractère de la relation entre $y$ et $x$ sera représenté sous la forme:

$$\left.\begin{array}{ll} y = f(x) & \text{pour } 0 < x < x_{1,n}; \\ y = \varphi_1(f(\psi_1(x))) & \text{pour } x_{2,n} < x < x_{1,n}; \\ y = \varphi_2(f(\psi_2(x))) & \text{pour } x_{2,n} < x < x_{3,n} \end{array}\right\} \quad (5)$$

etc. jusqu'un renversement de l'itération. Les formules (5) présentent une généralisation utile des formules (4). Ces dernières peuvent être obtenues des formules (5) dans ce cas particulier

$$\psi(x) = \psi_1(x) = x. \quad (6)$$

Enfin, nous obtiendrons le degré suivant de généralisation, en substituant la définition de $\varphi_n$ et de $\psi_n$ ($n = 1, 2, 3 \ldots$) résultant de (3) par une autre définition, plus générale, à savoir:

$$\left.\begin{array}{l} \varphi_1(x) = \varphi(x, x_{1,n}, 1); \\ \varphi_2(x) = \varphi[\varphi(x, x_{1,n}, 1), x_{2,n}, 2]; \\ \varphi_3(x) = \varphi\{\varphi[\varphi(x, x_{1,n}, 1), x_{2,n}, 2], x_{3,n}, 3\} \end{array}\right\} \quad (7)$$

etc. Faisons de même pour $\psi$. Les explications relatives à ((5) — (3)) et ((5) — (7)) sont données plus bas.

Établissons maintenant un domaine où les termes ((5) — (3)) et ((5) — (7)) proposés, et le procédé de Boltzman-Volterra seraient semblables.

§ 2. Écrivons l'équation de l'action ultérieure de Boltzman-Volterra sous la forme générale:

$$y(t) = \Phi\{x(t), \int_{-\infty}^{t} F_1[M(t-\tau), (x(\tau)]\, d\tau, \int_{-\infty}^{t}\int_{-\infty}^{t_1} F_2[\ ]\, d\tau\, d\tau_1 \ldots \quad (8)$$

où $t$ désigne le temps et dont le cas particulier est l'équation ordinaire, linéaire dans trois sens, de Boltzman-Volterra:

$$y(t) = M; \quad x(t) + \int_{-\infty}^{t} M(t-\tau)\, x(\tau)\, d\tau. \quad (8a)$$

[Linéaire dans trois sens exprime ce fait que: a) la fonction dépend *linéairement* de $x(t)$; b) n'a lieu que le fonctionnel *linéaire* $\int_{-\infty}^{t_1} M(t-\tau)\, x(\tau)\, d\tau$ et c) il rentre aussi dans l'expression de $y(t)$ *linéairement*].

Considérons d'abord le caractère de la solution (8), si nous nous bornons à l'étude d'un groupe des phénomènes, ayant la même vitesse, mais une amplitude différente (la terminologie sera expliquée par ce qui suit).

Déplaçons-nous d'abord de façon à accroître $x = x_1(t)$ de 0 à $x'_{1n}$. La fonction $y$ prendra, d'après (8), toute une série de valeurs $y = y_1(t)$. Si l'on exclut $t$, l'on aura:

$$y = f(x). \quad (1)$$

Si l'on considère une autre expérience, où $x = x_2(t)$ augmentait avec la même vitesse de $x$ à $x''_{1n}$ ($x''_{1n} > x'_{1n}$), nous obtiendrons également certaines $y = y_2(t)$. La coïncidence $x_1 t$ et $x_2(t)$, jusqu'une valeur égale à

$$x_1(t) = x_2(t) = x'_{1n}, \quad (9)$$

sera appelée par nous vitesse égale des processus, et la distinction de $x'_{1n}$ et $x''_{2n}$ — la différence des amplitudes. Il est évident, que nous obtiendrons

alors, en excluant $t$ de $x=x_2(t)$ et de $y=y_2(t)$, une fonction $y=f(x)$ coïncidant avec (1) jusqu'à la valeur $x=x'_{1n}$.

Comparons maintenant le caractère du mouvement inverse, en bas, de $x'_{1n}$ et de $x''_{1n}$, respectivement. Nous obtiendrons facilement

$$y=f_{1,1}(x) \text{ en premier cas;} \quad y=f_{1,11}(x) \text{ en second cas.}$$

Ce résultat peut être généralisé, si l'on écrit:

$$y=f_1(x, x_{1n}) \tag{10a}$$

où $x_{1n}$ résulte de $x'_{1n}$, ou bien de $x''_{1n}$, ou d'une autre valeur de $x$ dans le lieu de renversement du processus.

On trouve par un raisonnement analogue:

$$\left.\begin{aligned} y&=f_2(x, x_{1n}, x_{2n}) \text{ pour } x_{2n}<x<x_{3n} \\ y&=f_3(x_1, x_{1n}, x_{2n}, x_{3n}) \text{ pour } x_{4n}<x<x_{3n} \end{aligned}\right\} \tag{10}$$

etc., ce qui n'exige pas d'explications. Mais tout ceci n'a lieu que pour les processus ayant la même vitesse [dans le sens de (9)], mais aux amplitudes arbitraires ($x_{k,n}$, $-k=1, 2, 3\ldots$).

Quand aux processus aux vitesses différentes, les valeurs de $f_1$, $f_2$, $f_3$ peuvent être distinctes, même pour le phénomène ou la substance analogues. Si nous écrivons la condition

$$f_1(x, x_{1n})=\varphi_1[f(\psi_1(x, x_{1n})), x_{1n}] \tag{11a}$$

où $f_1$ et $f$ seront considérées comme fonctions connues, nous verrons que les valeurs de $\varphi_1$ et de $\psi_1$ pourront être définies par un nombre infiniment grand de procédés (en particulier on peut poser $\varphi_1=x$ ou $\psi=x$).

Puis des conditions

$$\left.\begin{aligned} f_2(x, x_{1n}, x_{2n})&=\varphi_2[f_1(\psi_2(x, x_{2n})\, x_{1n})\, x_{2n}] \\ f_3(x, x_{1n}, x_{2n}, x_{3n})&=\varphi_3[f_2(\psi_3(x, x_{3n})\, x_{1n}, x_{2n})\, x_{3n}] \end{aligned}\right\} \tag{11b}$$

etc. on détermine les fonctions

$$\varphi_2, \varphi_3, \ldots, \psi_2, \psi_3. \tag{12}$$

Ici on ne peut plus considérer, que l'une des grandeurs $\varphi_i$ et $\psi_i$ des couples soit arbitraire. Au contraire, à la première vue on pourrait croire que l'expression (11b) ne fut pas assez générale, et que soient possibles les cas où l'on puisse ne pas trouver $\varphi_i$ et $\psi_i$ de la forme correspondante à (11b), c'est-à-dire, qu'il y ait une dépendance explicite des $\varphi_2$ et $\psi_2$ de $x_{1n}$ des $\varphi_3$, et $\psi_3$ de $x_{1n}$ et $x_{2n}$, etc. Mais il est facile à démontrer que pour les conditions physiques évidentes: a) la continuité des fonctions $f_i$, et b) l'égalité des deux fonctions aux bornes des intervalles

$$f(x_{1n})=f_1(x_{1n}, x_{1n}); \quad f_1(x_{2n}, x_{1n})=f_2(x_{2n}, x_{1n}, x_{2n}) \tag{13}$$

etc., les équations (11b) ont leurs solutions sous la forme que nous voyons représentée par ces équations mêmes.

En effet, considérons, par exemple, la première des équations (11b). Soit d'abord $\psi_1=x$ et soit $(\underline{x}, \underline{x}_{1n})$ et $(\underline{\underline{x}}, \underline{\underline{x}}_{2n})$ deux couples de valeur telles que

$$f_1(\underline{x}, \underline{x}_{1n})=f_1(\underline{\underline{x}}, \underline{\underline{x}}_{1n}). \tag{A}$$

Alors, pour $x_{2n}$ donné on a

$$\varphi_1[f_1(\underline{x}, \underline{x}_{1n}), x_{2n}]=\varphi_1[f_1(\underline{\underline{x}}, \underline{\underline{x}}_{1n}, \underline{\underline{x}}_{2n})\, x_{2n}]. \tag{B}$$

Mais il ne suit de rien que l'on aurait

$$f_2(\underline{x}, \underline{x}_{1n}, x_{2n})=f_2(\underline{\underline{x}}, \underline{\underline{x}}_{2n}, x_{2n}); \tag{C}$$

bien au contraire, on pourrait s'attendre à voir l'inverse [c'est-à-dire que

la variante primitive de la méthode d'itération ne peut pas donner les équations analogues à (11b) et ne peut pas, par conséquent, être mise en concordance avec la méthode de Boltzman-Volterra].

Mais cela peut être corrigé par la substitution de $\psi(\bar{x}, x_{2n}) = \underline{x}$ et $\psi(\bar{\bar{x}}, x_{2n}) = \underline{\underline{x}}$, grâce à quoi les conditions (A) et (B) peuvent être satisfaites. Le choix des $\bar{x}$ et $\bar{\bar{x}}$ assurera la correspondance avec les variations de (C): $f_2(\bar{x}, \underline{x}_{1n}, x_{2n}) = f_2(\bar{\bar{x}}, \underline{\underline{x}}_{1n}, x_{2n})$.

Les fonctions $\varphi_1, \varphi_2, \varphi_3 \ldots$ peuvent être considérées, dans la suite, comme les fonctions particulières de l'ensemble

$$\varphi(x, x_{kn}, k) \tag{14}$$

$$\varphi_1 = \varphi(x, x_{1n}, 1); \quad \varphi_2 = \varphi(x, x_{2n}, 2) \tag{14a}$$

etc. De même on peut réduire $\psi_1, \psi_2, \psi_3$ à la forme quelconque

$$\psi(x, x_{k,n}, k). \tag{15}$$

[La troisième variable de (14) et de (15) ne prend que les valeurs entières, uniquement pour les raisons de simplicité d'écriture. En principe, on peut supposer dans (14) et (15) une série quelconque de valeurs du paramètre $k$].

Par conséquent, la solution la plus générale de relation Boltzman-Volterra se trouve réduite à la forme de la méthode généralisée d'itération ((5)—(7)). Il est vrai que cela n'a lieu que pour les processus à l'amplitude différente, mais ayant des mêmes vitesses. Il est aussi évident, que la méthode d'itération sous la forme ((5)—(7)) est entièrement équivalente, au point de vue mathématique, à celle de Boltzman-Volterra pour les processus, où le phénomène de l'action ultérieure (hystéresis) ne dépend pas du tout (ou bien pratiquement) de la vitesse.

*Ainsi pour les restrictions citées, de même que pour d'autres, évidentes* (a) *et* (b), *indiquées plus haut, une multiplicité de fonctions physiquement réelles, satisfaisant à la forme de description de Boltzman-Volterra, et une multiplicité de fonctions correspondant à la forme de description* ((5)—7)), *coïncident l'une avec l'autre. Par conséquent, pour de tels cas* (*les phénomènes indépendants de la vitesse, de même que les phénomènes à la même vitesse, mais aux amplitudes différentes*) *nous pouvons, sans perdre de sens physique, remplacer la description de Boltzman-Volterra par la description* ((5)—(7)), *qui est incomparablement plus simple.*

Mais tout cela ne nous donne pas encore le droit d'affirmer, que le cas exposé soit la seule correspondance possible. Cela fera objet d'une autre étude. Maintenant nous indiquerons, en terminant, les points essentiels de la généralisation proposée de la méthode d'itération servant à décrire l'effet postérieur et l'hystéresis, à savoir:

a) l'introduction de l'itération «intérieure» ($\psi$), sans quoi la correspondance avec la méthode de Boltzman-Volterra est impossible [voir les formules (A), (B) et (C)];

b) la généralisation de l'affirmation sur «l'identité» des itérations en passant de l'identité mathématique à l'uniformité physique, ce qui est supposé dans (7) et ce qui doit effectivement avoir lieu, comme cela résulte des raisonnements élémentaires. Dans la suite nous allons le prouver en établissant la forme concrète de $\varphi$ et $\psi$.

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*PHYSIK*

# ZUR THEORIE DES SPINS ELEMENTARER TEILCHEN

Von V. L. GINSBURG

*(Vorgelegt von S. I. Wawilow, Mitglied der Akademie, den 25. II. 1941)*

Die Unzulänglichkeit der bestehenden Theorie der Wechselwirkung der Mesotronen mit dem elektromagnetischen Feld und mit schweren Teilchen äussert sich unter anderem in der dem Experiment und einigen allgemeinen Forderungen widersprechenden Energieabhängigkeit der Wirkungsquerschnitte für die Streuung von Licht an einem Mesotron und der Mesotronen am Proton-Neutron. Die sinnwidrigen Resultate, zu denen man gelangt, können im letzgenannten Falle als Auswirkungen der Trägheitslosigkeit des Spins und der Ladungsvertauschung des schweren Teilchens gedeutet werden ([1]); im Falle der Streuung von Licht am Mesotron ist die Zunahme der Wirkungsquerschnitte eine Folge der Trägheitslosigkeit des Mesotronspins ([2]). Wie Heisenberg nachgewiesen hat, findet die unbegrenzte Zunahme der Wirkungsquerschnitte für die Streuung von Licht am magnetischen Moment auch in der klassischen Theorie statt, sofern man von der Rückwirkung des Eigenfeldes dieses Moments auf seine Bewegung absieht; wird diese Rückwirkung jedoch berücksichtigt, so nimmt der Wirkungsquerschnitt für hohe Frequenzen einen konstanten Wert an. Heisenberg ([1]) hat sich nur auf ein konkretes Beispiel beschränkt, weshalb in seiner Untersuchung einige interessante Besonderheiten der Rückwirkung des magnetischen Moments auf sich selbst unaufgeklärt geblieben sind; wie mir scheint, dürfte das richtige Verständnis dieser Besonderheiten einen wichtigen Fortschritt in dem gesamten Spinproblem ermöglichen.

Zur Klärung dieser Besonderheiten untersuchen wir die Bewegung eines ladungsfreien magnetisierten Teilchens, dessen Mittelpunkt ruht. Es sei die Magnetisierung des Teilchens $\vec{M} = \vec{\mu} D(\vec{r})$, wo $\vec{\mu}$ das magnetische Moment und $D(\vec{r})$ den Formfaktor bedeuten $\left(\int D(\vec{r})\, dv = 1\right)$. Das System der Gleichungen der Elektrodynamik und Mechanik erhält für unseren Fall die Form

$$\left.\begin{aligned} &\frac{d\vec{s}}{dt} = \int [\vec{\mu} H]\, D(\vec{r})\, d\sigma + [\vec{\mu}\vec{H}_e]; \\ &\Box \vec{A} = -4\pi \operatorname{rot} \vec{M}; \quad \vec{H} = \operatorname{rot} \vec{A}; \quad \vec{H}_e - \text{äusseres Feld.} \end{aligned}\right\} \qquad (1)$$

Das System (1) kann durch die Beziehung zwischen dem Drehimpuls $\vec{s}$ und $\vec{\mu}$ ergänzt werden; wir setzen, wie üblich, $\vec{\mu} = \varkappa \vec{s}$. Hier besteht die Möglichkeit, im allgemeinen $\vec{H}$ aus (1) zu eliminieren, wobei sich für

$\vec{s}$ die nachstehende Bewegungsgleichung ergibt, die die Rückwirkung des Eigenfeldes berücksichtigt (²):

$$\left.\begin{array}{l}\dot{\vec{s}} = \frac{d\vec{s}}{dt} = \varkappa[\vec{s}\vec{H}_e] - \frac{4\nu_m \varkappa^2}{3\pi c^3}[\vec{s}\dot{\vec{s}}] + \frac{2\varkappa^2}{3c^3}[\dot{\vec{s}}\dot{\vec{s}}] + \text{Glieder, die proportional} \\ \frac{1}{\nu_m^n}(n = 1, 2, 3 \ldots) \text{ sind, } + \text{Glieder, die bei } t \gg \frac{1}{\nu_m} \text{ verschwindend klein werden.}\end{array}\right\} \quad (2)$$

In Gleichung (2) bedeutet $\nu_m$ diejenige Frequenz, bei der das Spektrum der Fourier-Komponenten des Feldes abreisst; offensichtlich ist $\nu_m \sim \frac{c}{r_0}$, wo $r_0$ den Radius des Teilchens darstellt. Man beachte, dass in (2) ein reaktives Glied der Form $\frac{d\vec{\mu}}{dt}$ fehlt, das erst im Falle geladener Teilchen auftritt. Das Glied $\vec{R} = \frac{2\varkappa^2}{3c^3}[\dot{\vec{s}}\ddot{\vec{s}}]$ in (2) ist dissipativ und der Kraft der Strahlungsreaktion $\frac{2e^2}{3c^3}\dddot{x}$ analog. Das Moment $\vec{L} = -\frac{4\nu_m \varkappa^2}{3\pi c^3}[\dot{\vec{s}}\ddot{\vec{s}}]$ bedeutend grösser ist, als $\vec{R}\left(\text{wenn } H_e \sim \cos \nu t, \text{ wird } \frac{R}{L} \sim \frac{\nu}{\nu_m}\right)$, die Konservativität der Gleichung für $\vec{s}$ nicht beunrechtigt und bedingt eine *wesentliche Änderung der Dynamik des «Spins» eines Teilchens.* Vernachlässigt man bei der Streuung von Licht am Moment $\vec{\mu}$ das Glied $\vec{L}$, d. h. beschränkt man sich auf die Gleichung $\frac{d\vec{s}}{dt} = \varkappa[\vec{s}\vec{H}_e]$, so wird der Wirkungsquerschnitt (¹):

$$\sigma = \frac{8\pi\mu_0^2\varkappa^2\nu^2}{3c^4}; \quad \mu_0^2 = \mu_x^2 + \mu_y^2; \quad H = H_z. \quad (3)$$

Der nämliche Wirkungsquerschnitt ergibt sich auch aus der Pauli'schen Quantentheorie, wo $\mu_0 = \frac{e\hbar}{2mc}$.

Das Moment $L$ wird bei $\lambda = \frac{c}{\nu} \sim \frac{\hbar}{mc}$ seiner Grössenordnung nach gleich $\frac{ds}{dt}$, wenn man $\varkappa$ den Wert $\frac{e}{mc}$ zuschreibt und ferner ansetzt $\nu_m \sim \frac{c}{r_0} \sim \sim \frac{mc^3}{e^2}$ und $\mu \sim \varkappa\hbar$; bei denselben Voraussetzungen erhält man $R \sim \frac{ds}{dt}$, wenn $\lambda \sim \frac{e^2}{mc^2} \sim r_0$. Ist $\lambda \gg \frac{\hbar}{mc}$, so darf das Glied $\frac{ds}{dt}$ gegen $L$ vernachlässigt werden, und man erhält, wie leicht ersichtlich

$$\sigma = \frac{3\pi^3 c^2}{2\nu_m^2} \sim r_0^2. \quad (4)$$

Der Charakter der Spinbewegung, gegeben durch den Ausdruck

$$\frac{d\vec{s}}{dt} = \varkappa[\vec{s}\vec{H}_e] - \frac{4\nu_m\varkappa^2}{3\pi c^3}[\vec{s}\ddot{\vec{s}}], \quad (5)$$

kann klarer erfasst werden, sofern man diesen Ausdruck durch äquivalente Systeme von Gleichungen erster Ordnung ersetzt:

$$\frac{d\vec{\sigma}}{dt} = \varkappa[\vec{s}\vec{H}_e]; \quad \frac{d\vec{s}}{dt} = -\alpha[\vec{s}\vec{\sigma}]; \quad \alpha = \frac{3\pi c^3}{4\nu_m\varkappa^2 s^2}; \; s^2 = \text{const} \quad (6)$$

oder

$$\frac{d\vec{\rho}}{dt} = \varkappa[\vec{s}\vec{H}_e] + \alpha[\vec{s}\vec{\rho}]; \quad \frac{d\vec{s}}{dt} = -\alpha[\vec{s}\vec{\rho}]; \quad \vec{\rho} = \vec{\sigma} - \vec{s}. \quad (6')$$

Durch Eliminierung von $\vec{\sigma}$ aus (6) oder von $\vec{\rho}$ (6′) gelangt man zu (5). In dem gewöhnlichen Schema der Quantentheorie der Strahlung wird das Eigenfeld des Teilchens nicht berücksichtigt*; im Falle eines magnetischen Teilchens bedeutet dies, dass man das Quantenanalogon der Gleichung (5), jedoch ohne das Glied $\vec{L}$, verwendet. Dadurch wird es verständlich, dass hier eine Zunahme der verschiedenen effektiven Querschnitte mit der Energie stattfindet. Die Berücksichtigung des Eigenfeldes unter Zuhilfnahme der Methoden der Strahlungstheorie erscheint nahezu hoffnungslos. Dementgegen darf es als ein naturgemässes Bestreben gelten, *den Spinbegriff in der Quantentheorie dahin zu bereichern, damit er dem klassischen, durch Gleichung* (5) *bestimmten Spin entspricht.* Es sei daher $\vec{s}$ in Gleichung (5) ein Operator; in derselben Gleichung ersetzen wir aus Symmetriegründen $[\vec{s}\,\vec{\dot{s}}]$ durch $\frac{1}{2}\{[\vec{s}\,\vec{\dot{s}}]-[\vec{\dot{s}}\,\vec{s}]\}$. Alsdann lässt sich Gleichung (5) mit der vorgenommenen Substitution aus dem Hamiltonschen Operator

$$H=\beta(\vec{s}\vec{\rho})-\varkappa(\vec{s}\vec{H}_e) \tag{7}$$

herleiten, sofern

$$\left.\begin{array}{l}[\vec{\rho}\vec{\rho}]=i\hbar\vec{\rho},\ \rho_i s_k=s_k\rho_i, \quad \beta=\dfrac{3\pi c^3}{4\nu_m\varkappa^2\left(2s^2-\dfrac{\hbar^2}{2}\right)}\\ [\vec{s}\vec{s}]=i\hbar\vec{s},\ \ s_i s_k+s_k s_i=0\,(i\neq k),\ \ s_1^2=s_2^2=\\ =s_3^2,\ s^2=s_1^2+s_2^2+s_3^2=\text{const (Zahl)}.\end{array}\right\} \tag{8}$$

Somit wird die Eigendrehung des Teilchens auch in der Quantentheorie durch zwei Momente $\vec{s}$ und $\vec{\rho}$ beschrieben. Aus (7) — (8) ergeben sich folgende Erhaltungssätze ($\vec{\sigma}=\vec{s}+\vec{\rho}$):

$$\vec{s}^2=\text{const},\ \ \vec{\rho}^2=\text{const},\ \vec{\sigma}\vec{H}_e=\text{const},\ \ \vec{\sigma}^2-\frac{2\varkappa}{\beta}(\vec{s}\vec{H}_e)=\text{const}. \tag{9}$$

Ohne Feld bleibt auch der Vektor $\vec{\sigma}$ erhalten, dem hier die Rolle des vollen Teilchenmoments zukommt. Als Beispiel behandeln wir den einfachsten Sonderfall, wo $\vec{\rho}$ und $\vec{s}$ in Form vierreihiger Matrizen dargestellt werden können, die man aus der Diracschen Theorie kennt:

$$\left.\begin{array}{l}s_1=\frac{\hbar}{2}\begin{pmatrix}0&I\\I&0\end{pmatrix};\quad s_2=\frac{\hbar}{2}\begin{pmatrix}0&-iI\\iI&0\end{pmatrix};\quad s_3=\frac{\hbar}{2}\begin{pmatrix}I&0\\0&-I\end{pmatrix};\quad I=\begin{pmatrix}1&0\\0&1\end{pmatrix}\\ \rho_i=\frac{\hbar}{2}\begin{pmatrix}\tau_i&0\\0&\tau_i\end{pmatrix};\quad \tau_1=\begin{pmatrix}0&1\\1&0\end{pmatrix};\quad \tau_2=\begin{pmatrix}0&-i\\i&0\end{pmatrix};\quad \tau_3=\begin{pmatrix}1&0\\0&-1\end{pmatrix}\end{array}\right\} \tag{10}$$

Die Eigenwerte der Matrix $\beta$ $(\vec{s}\vec{\rho})$ sind in diesem Falle $E_1=E_2=E_3=\frac{\hbar^2}{4}\beta$, $E_4=-\frac{3}{4}\hbar^2\beta$; die Eigenwerte von $\sigma^2$ sind $2\hbar^2$, $2\hbar^2$, $2\hbar^2$, 0, und die Eigenwerte von $\sigma_i$ sind $0, \pm\hbar, 0$. Somit beträgt das volle Moment des Teilchens entweder 0 oder $\sqrt{2}\hbar$. Die Zustände mit $\sigma^2=2\hbar^2$ sind die oberen (angeregten) Zustände, da $\beta>0$. Bei Vorhandensein des Feldes $\vec{H}_e$ sind zwischen allen Zuständen Übergänge möglich. Die Eigenfunktionen der störungsfreien Aufgabe ($H=\beta(\vec{s}\vec{\rho})$) sind:

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* Dabei bleibt im Falle eines magnetischen Teilchens ein weitaus grösserer Teil seiner Rückwirkung unberücksichtigt, als im Falle eines Teilchens ohne magnetisches Moment (²).

$$\psi_{E_1}=\begin{pmatrix}1\\0\\0\\0\end{pmatrix};\quad \psi_{E_2}=\begin{pmatrix}0\\0\\0\\1\end{pmatrix};\quad \psi_{E_3}=\frac{1}{\sqrt{2}}\begin{pmatrix}0\\1\\1\\0\end{pmatrix};\quad \psi_{E_4}=\frac{1}{\sqrt{2}}\begin{pmatrix}0\\1\\-1\\0\end{pmatrix}. \tag{11}$$

Definiert man in (7) $\vec{H}_e$ als den magnetischen Vektor des Strahlungsfeldes, so kann man die Streuung von Licht an dem gegebenen Teilchen berechnen. Wenn $\hbar\nu < \Delta = E_1 - E_4$, so ist nur elastische Streuung möglich (die Bewegung des Teilchens wird nicht berücksichtigt). Bei $\hbar\nu \ll \Delta$ ist der Wirkungsquerschnitt gleich dem Ausdruck (3), multipliziert mit $\left(\frac{\hbar\nu}{\Delta}\right)^2$. Die Abnahme des Wirkungsquerschnitts hängt damit zusammen, dass im Grundzustand $E_4$ der Spin Null ist. Bei $\hbar\nu > \Delta$ wird unelastische Streuung möglich, für die der Wirkungsquerschnitt bei $\hbar\nu \gg \Delta$ gegen Null strebt; für elastische Streuung strebt der Wirkungsquerschnitt dabei gegen (4), d. h. er wird abgeschnitten.

Im allgemeinen Falle, d. h. wenn man sich nicht auf die Vorstellung der vierreihigen Matrizen beschränkt, kann das Teilchen eine beliebige Anzahl angeregter Zustände und einen beliebigen Spin besitzen. Wir wollen weder den sich hier eröffnenden Möglichkeiten nachgehen, noch den Einfluss der Bewegung des Teilchenschwerpunktes untersuchen, da die entwickelte Theorie unrelativistisch und folglich offenbar unzureichend ist. Die gewonnenen Ergebnisse sind jedoch auch für die relativistische Theorie der Elementarteilchen von Interesse. In der Tat besitzt das magnetische Teilchen, wie oben erwiesen, Innenenergie*, die in die Hamilton-Funktion oder in den Hamilton-Operator eingeht [das Glied $\beta(\vec{s}\vec{\rho})$ in Gleichung (7)]. Offensichtlich müssen in der relativistischen klassischen Theorie des magnetischen Moments bei der Berücksichtigung seines Eigenfeldes ebenfalls solche Glieder vorhanden sein, und für die Hamilton-Funktion gilt daher nicht mehr

$$H = \frac{E^2}{c^2} - p^2 = m^2c^2. \tag{12}$$

Daher muss, allgemein gesehen, aus der Wellengleichung des relativistischen Teilchens (bei Nichtvorhandensein des Feldes) nicht die klassische Theorie mit Gleichung (12) hervorgehen. Indessen beruht jedoch auf dieser Annahme die ganze Theorie der Teilchen mit willkürlichem Spin wie sie von Dirac, Pauli und Fierz (4) (im weiteren D-P-F) entwickelt worden ist; dasselbe bezieht sich auch auf die Diracsche Gleichung für das Elektron. Im letztgenannten Falle ist die obige Annahme gerechtfertigt, was mit dem Fehlen eines «echten» magnetischen Moments beim Elektron zusammenhängt (5,6). Für das Mesotron, das ein solches Moment besitzt, trifft dies schon nicht mehr zu. Daher erscheint uns die *Aufgabe aktuell, Wellengleichungen zu ermitteln, deren Lösungen bei Nichtvorhandensein von Feldern der Klein-Gordonischen Gleichung nicht genügen.* Auf diesem Wege eröffnet sich eine Reihe neuer Perspektiven, während ohne Verzicht auf die Annahme D-P-F ein weiterer Fortschritt wohl kaum möglich ist.

Bei der Lösung der gestellten Aufgabe kann man versuchen, die klassische Hamilton-Funktion zu ermitteln, die im Falle des magnetischen Teilchens die Gleichung (12) zu ersetzen hat. Vernachlässigt man die Beziehung zwischen Spin und Bahn (wofür ich keine besonderen Gründe zu erblicken vermag), so lässt sich die betreffende Formel leicht ermit-

* Im klassischen Falle ist diese Energie mit der kinetischen Umdrehungsenergie verbunden.

teln, indem man die Gleichungen (6′) bei Nichtvorhandensein des Feldes relativistisch verallgemeinert und die relativistische Kramerssche Theorie (7) in Anspruch nimmt. Somit erhält man:

$$H = \frac{E^2}{c^2} - p^2 - 2m\alpha\,(\vec{s}\vec{\rho}) = m^2c^2 \tag{13}$$

wo

$$\{s_1 s_2\} = -s_3, \quad \{\rho_1\rho_2\} = -\rho_3, \quad \{s_i\rho_k\} = 0 \text{ usw.},$$

wobei $\vec{s}$ und $\vec{\rho}$ Grössen darstellen, die in Hinsicht ihrer Transformationseigenschaften dem Ausdruck $\vec{H} + i\vec{E}$ ähnlich sind ($\vec{H}$ and $\vec{E}$ sind die Feldstärken des magnetischen, bzw. elektrischen Feldes; { } — die Poissonschen Klammern). Die Funktion (13) ist zur Ermittelung relativistischer Wellengleichungen noch nicht benutzt worden. Es handelt sich darum, dass die relativistischen Invarianzforderungen derart starr sind, dass es vielleicht leichter sein dürfte, die den Bedingungen D-P-F nicht genügenden Wellengleichungen unmittelbar zu suchen. Als Beispiel sei hier darauf hingewiesen, dass man Gleichungen solcher Art aus der Lagrange-Funktion erhalten kann:

$$\left.\begin{aligned} L = &- aU_\mu U^*_\mu - \frac{1}{2} G_{\mu\nu} G^*_{\mu\nu} - bV_\mu V^*_\mu - \\ &- \frac{1}{2} F_{\mu\nu} F^*_{\mu\nu} - \alpha(U_\mu V^*_\mu + \text{k.k.}) - \\ &- \frac{\beta}{2}(G_{\mu\nu} F^*_{\mu\nu} + \text{k.k.}); \quad a \neq b \end{aligned}\right\} \tag{14}$$

wo k.k. — komplex konjugierte Funktion, $U_\mu$ und $V_\mu$ — Komponenten der vierdimensionalen Vektoren darstellen [Bezeichnungen s. bei Bhabha (8); $G_{\mu\nu}$ erhält man aus $U_\mu$, $F_{\mu\nu}$ aus $V_\mu$]. Ist $\alpha = \beta = 0$, so führt (14) zu zwei Gleichungen für unabhängige Mesotronen mit verschiedener Masse; ist dagegen $\alpha$ oder $\beta$ nicht Null, so gelangt man zu einer Wellengleichung für ein Teilchen mit dem Spin Eins und zwei Massewerten. Zwischen den Zuständen mit verschiedenen Massen sind unter der Einwirkung des Feldes Übergänge möglich. Die hier angeführte Gleichung ist mit der Spintheorie nicht unmittelbar verbunden und wurde lediglich zwecks Illustration gebracht. Die Gleichung zeigt ferner, dass der Verzicht auf die Bedingung D-P-F nicht unbedingt mit der Berücksichtigung der Selbstenergie des Spins verknüpft sein muss, sondern dass dieser Verzicht auch bei Vorhandensein anderer innerer Freiheitsgrade des Teilchens notwendig ist. Zur Ermittelung des Zusammenhanges mit der vorhin dargelegten unrelativistischen Theorie empfiehlt sich naturgemäss die Analyse von Gleichungen, hergeleitet aus der Lagrange-Funktion, welche das Mesotron mit einem Teilchen mit dem Spin Null verbindet, usw. Gleichungen solchen Typus werden zurzeit untersucht.

Zum Schlusse spreche ich Prof. Ig. Tamm meinen Dank für seine Ratschläge und Diskussionen aus.

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*PHYSICS*

# ULTRASONIC WAVE VELOCITY IN AQUEOUS MIXTURES OF SOME ORGANIC LIQUIDS

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It is a well-known fact that some binary liquid mixtures show a maximum velocity and maximum coefficient of ultrasonic wave absorption at a given concentration. These deviations were discovered in acetone—water [1] and ethyl alcohol—water [2] mixtures. In both mixtures the ultrasonic velocity maximum may be observed at approximately a 30% volume concentration of the alcohol or acetone and the absorption coefficient maximum at a concentration of 60%.*

Another feature characteristic of both these mixtures is the viscosity maximum. Such departures from the additivity law point to the existence of considerable molecular interaction between the components of the mixture. Thus, it is known that the presence of a viscosity maximum serves as the sign of molecular linkage in liquid mixtures, i. e. of the formation of complex molecules as result either of physical association or of chemical combination.

It is interesting at this point to mention that, as P. A. Bažulin and J. M. Merson have pointed out, the calculation of the absorption coefficient for the mixture acetone—water, according to Stokes' formula, taking into account only the usual viscosity coefficient, fixes the maximum also at 60%. It is therefore quite possible that the increase in the absorption of ultrasonic waves at this concentration is connected with the physical and chemical phenomena just spoken of. The ultrasonic velocity maximum in these mixtures occurs in a region, which corresponds to lower concentrations than does the viscosity maximum. Furthermore, the velocity of sound is determined by the expression $v = \sqrt{\frac{1}{\beta_{ad}\rho}}$, which is independent of viscosity. However, the existence simultaneously of both maxima in these mixtures is noteworthy. A similar regularity permitted the author of the present paper to claim [2] that similar phenomena may be expected to take place in other mixtures. Further measurements, carried out by the author, of ultrasonic wave velocity in a number of mixtures proved this indeed to be the case. The results of these measurements in a number of binary mixtures, possessing viscosity maxima at different concentrations, according to the data of several authors, are given below.

The measurement of sound velocity was achieved by photographing diffraction spectra due to the grating generated by ultrasonic waves. The negatives were taken on film in monochromatic light with a «FED» camera. The distance between spectra of the $\pm$ 1-order was measured by means of a comparator. The constant of the device was determined by measuring the photographs taken of diffraction spectra produced by a wire grating (the grating constant was 0.702 mm). To minimize the effect of temperature variations due to the piezo-quartz, a glass cube (3 ×3 ×3 cm) was introduced

* The author of this paper was able to detect qualitatively the presence of an absorption maximum in the mixture ethanol—water. The author is indebted to J. M. Merson for having informed us later that the measurement of absorption in this mixture showed a maximum at a volume concentration of alcohol of 60%.

| % by volume | Methyl alcohol—water $t=25°C$ | | Acetone—water $t=20.5°C$ | | Acetic acid—water $t=27°C$ | | Pyridine—water $t=23°C$ | | Aniline—acetic acid $t=15°C$ | |
|---|---|---|---|---|---|---|---|---|---|---|
| | % by weight | v, m/sec | % by weight | v, m/sec | % by weight | v, m/sec | % by weight | v, m/sec | % by weight | v, m/sec |
| 0 | 0 | — | 0 | 1484 | 0 | 1502 | 0 | 1490 | 0 | 1671 |
| 10 | 8.1 | 1531 | 8.1 | 1531 | 10.3 | 1527 | 9.8 | 1543 | 10.3 | 1625 |
| 20 | 16.5 | 1557 | 16.5 | 1563 | 20.6 | 1535 | 19.7 | 1560 | 20.5 | 1582 |
| 30 | 25.4 | 1569 | 25.1 | 1576 | 30.9 | 1525 | 29.6 | 1574 | 30 7 | 1555 |
| 40 | 34.6 | 1554 | 34.5 | 1558 | 41.0 | 1509 | 39.5 | 1581 | 40.8 | 1522 |
| 50 | — | — | — | — | 51.0 | 1484 | — | — | 50.8 | 1496 |
| 60 | 54.3 | 1476 | 54.3 | 1466 | 61.0 | 1451 | 59.5 | 1588 | 62.8 | 1461 |
| 70 | — | — | — | — | — | — | — | — | 70.7 | 1411 |
| 80 | 76.0 | 1313 | 76.0 | 1343 | 81.1 | 1353 | 79.7 | 1544 | 80.5 | 1351 |
| 90 | — | — | — | — | — | — | — | — | 90.3 | 1272 |
| 100 | 100 | 1113 | 100 | 1134 | 100.0 | 1158 | 100 | 1436 | 100.0 | 1194 |

between the piezo-quartz and the next cell containing the liquid. The measurements were conducted at frequencies of 5596.3 and 4967.5 kHz. In the table are given the results obtained in measuring sound velocity in all the mixtures investigated by the author.* Figs. 1, 2, 3 and 4 represent curves of sound velocity, density and the coefficient of adiabatic compressibility as function of weight concentration for 4 mixtures: methyl alcohol—water, acetic acid—water, pyridine—water and aniline—acetic acid. The coefficient $\beta_{ad}$ was calculated from sound velocity data with the aid of standard tables of densities.

The mixtures methyl alcohol—water (Fig. 1), ethyl alcohol—water and acetone—water must, according to the results obtained, be grouped as mixtures of the same type. It is characteristic of these mixtures that density is almost linearly dependent on concentration, that sound velocity has a maximum at approximately 30% for all the mixtures and that the coefficient $\beta_{ad}$ has a minimum at the same concentration. The viscosity maximum of all these mixtures is located in the concentration range of 40—50%. The mixture acetic acid—water (Fig. 2) shows a density maximum and a viscosity maximum at 80%. Maximum velocity and minimum $\beta_{ad}$ occur at 20% concentration by weight. In the mixture pyridine—water (Fig. 3) maximum velocity is observed at 60% and coincides approximately with maximum density and maximum viscosity. Finally, the mixture aniline—acetic acid shows a viscosity maximum ([4]) at 60% and a density maximum at about 70%. The velocity curve (Fig. 4) has no maximum, but at 60—70% displays a considerable inflexion towards increased values of velocity**.

These results in the first place afford evidence that it is only the aqueous mixtures that possess a velocity maximum. In the second place in all these mixtures it is the way in which $\beta_{ad}$—but not the density—changes, that determines the character of the sound velocity. This fact was pointed out some time ago by Parthasarathy ([3]), who reached this conclusion after measuring the velocity of sound in some non-aqueous mixtures. A striking example of this fact is, furthermore, offered by the mixture acetic acid—water, which has a velocity maximum and $\beta_{ad}$ minimum at 20%, although the density maximum lies at 80%.

* The mixture acetone—water was investigated by N. Prozoroff, according to P. A. Bažulin and J. M. Merson. The author also measured the velocity of sound in this mixture.

** It is probable that in all these mixtures an ultrasonic absorption coefficient maximum may also be observed.

The $\beta_{ad}$ minimum in the mixture pyridine—water and the inflection of the $\beta_{ad}$ curve towards decreasing values in the aniline—acetic acid mixture correspond approximately to the viscosity and density maximum. It is possible that these changes experienced by $\beta_{ad}$ are in some way connected with the above-mentioned physical and chemical phenomena. This, however, cannot be said of mixtures of the type ethyl alcohol—water and acetic acid—water. The presence in these mixtures of a $\beta_{ad}$ minimum situated in a region of a lower concentration, than the viscosity maximum, is, evidently, due

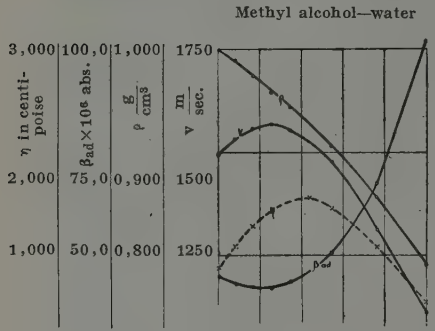


Fig. 1.

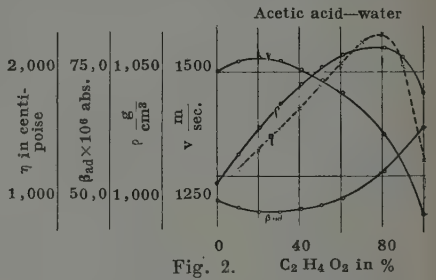


Fig. 2.

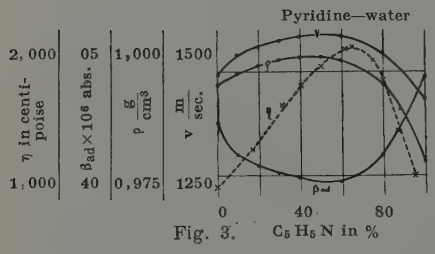


Fig. 3.

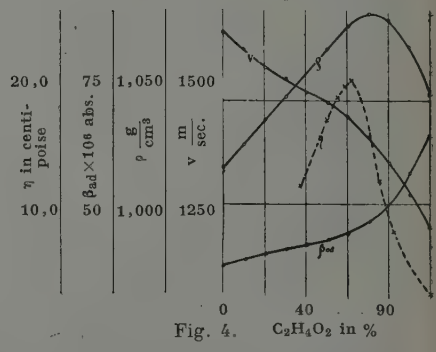


Fig. 4.

to totally different causes. Another circumstance inviting attention is the fact that in mixtures of the type ethyl alcohol—water a minimum of compressibility and a maximum of sound velocity occur at one and the same concentration, viz.~30% whatever the kind of mixture.

Thus, the nature of intromolecular interaction, which leads to such considerable changes in the elastic properties of mixtures of this type at low concentrations, is of special interest and calls for further study.

Lastly, it is interesting to note that according to X-ray investigations [4], the mixtures ethyl alcohol—water and pyridine—water belong to different structural groups, as shown by the nature of the variations in their roentgenograms with concentration in the constituentr of the mixture.

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22. II. 1941.

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*Translated by B. Belitzky.*

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*PHYSICS*

# CRITICAL FIELDS OF SUPERCONDUCTIVE VANADIUM

By N. E. ALEXEYEVSKY

*(Communicated by P. L. Kapitza, Member of the Academy, 20. II. 1941)*

As there are no data on the dependence of critical field values upon temperature for superconductive vanadium, it seemed of interest to make corresponding measurements*.

The specimen of vanadium kindly put at our disposal by P. L. Kapitza, Member of the Academy, was a plate about $0.1 \times 8 \times 4$ mm. The specimen

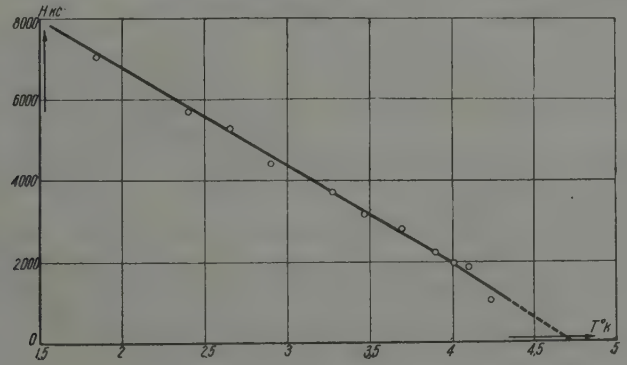


Fig. 1.

was attached to the end of a suspended device which consisted in a long glass rod, approximately 0.5 mm in diameter, fitted with a small mirror and suspended by a quartz string. The entire device was inserted into a glass tube filled with helium at a low pressure to create thermal contact. The tube was then placed into a Dewar vessel filled with liquid helium.

In the superconductive state when an external field was applied, the specimen orientated itself parallel to the field. The value of the field at which the specimen returned to the position of equilibrium, corresponded to the critical field value. The critical fields thus obtained for various temperatures are given in Fig. 1. It is evident from Fig. 1 that the investigated specimen of vanadium showed extremely high critical field values. The value

* The measurements were made at the Institute for Physical Problems in Moscow

of $\frac{dH_k}{dT}$ derived from the curve was 2400 Gauss per degree, which considerably exceeds all known values of $\frac{dH_k}{dT}$ for superconductive metals.

The value obtained by extrapolation for the critical temperature was 4.73° K, i. e. by 0.53° K greater than those observed by Meissner and Westerhoff ([1]). It should, however, be pointed out that the value obtained by extrapolation for the critical temperature may be somewhat overestimated. On the other hand, it seems quite probable that, since the specimen bore traces of rolling, the stresses could result in a somewhat increased value of the critical temperature. The critical values of the magnetic field, corresponding to the appearance of resistance, were measured with poor exactitude. However, on the basis of preliminary measurements it is possible to conclude that they are at any rate higher than values corresponding to the change in magnetic moment.

Ukrainian Physical-Technical Institute.
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*GEOPHYSIK*

# ÜBER DIE GRÖSSEN DER INTENSITÄT DER SONNENRADIATION IN DER ARKTIS

Von N. N. KALITIN

*(Vorgelegt von A. A. Grigoriev, Mitglied der Akademie, den 7. II. 1941)*

Die an der Erdoberfläche gemessene Grösse der Intensität der Sonnenradiation wird von der «Masse» der Atmosphäre, die von der Strahlenenergie der Sonne durchdrungen wird, d. h. von der Höhe der Sonne über dem Horizont, der Höhe des Beobachtungsortes über dem Meeresspiegel, der Entfernung der Erde von der Sonne und der Durchsichtigkeit der Atmosphäre, abhängen. Werden die Messungen immer bei einer und derselben «Masse» der Atmosphäre, sowie bei einer und derselben Höhe über dem Meeresspiegel vorgenommen, und auf die mittleren Entfernung der Erde von der Sonne bezogen, so werden die Variationen der gemessenen Grössen nur durch die Veränderung der Durchsichtigkeit der Atmosphäre bedingt.

Im gegebenen Falle wird vermutet, dass die Grösse der Sonnenkonstanten wirklich konstant bleibt, worauf wir bei der Genauigkeit der aktinometrischen Messungen heutzutage berechtigt sind. Gewöhnlich wird auch diese Methode der Aktinometrie beim Studium der Durchsichtigkeit der Atmosphäre angewandt.

Schon die ersten in den arktischen Verhältnissen ausgeführten systematischen Messungen der Intensität der Sonnenradiation zeigten, dass dort viel grössere Intensitäten beobachtet werden, als in südlicheren Orten; dieses wird durch kleineren Inhalt an Wasserdampf und kleinere Verstaubung der Atmosphäre in der Polargegend erklärt. Diese Tatsache wird auch durch weitere Beobachtungen bestätigt, die in der Tab. 1 in kurzer Zusammenfassung angeführt sind.

In der Tabelle sind die Beobachtungsorte nach abfallenden Breitengraden geordnet, wobei zum Vergleich auch Werte für vier weitere südliche Orte angeführt sind. Die angeführten Angaben sind Mittelwerte der Intensität der Sonnenradiation für den Mai und Juni, auf die mittlere Entfernung zwischen der Erde und der Sonne bezogen und für die «Masse» 2 gemessen, d. h. für einen Höhenstand der Sonne von 30°.

Wie aus der Tabelle zu ersehen ist, verkleinert sich mit der abfallenden geographischen Breite die Intensität der Sonnenradiation.

Die systematische Arbeit der aktinometrischen Station der Hauptverwaltung des nordischen Seewegs hat uns schon eine genügende Menge Material geliefert, welches die Möglichkeit gibt, die Jahresveränderungen der beobachteten Radiationswerte zu studieren und sie auch mit denen für südlichere Orte zu vergleichen.

Tabelle 1

Mittlere Intensität der Sonnenradiation bei einer Sonnenhöhe von 30°

| Ort | Geographische | | Mai, Kal. | Juni, Kal. |
|---|---|---|---|---|
| | Breite | Länge | | |
| Bucht Tichaja | 80°19′ | 52°48′ | 1.28 | 1.31 |
| Insel Ujedinenija | 77°30′ | 82°12′ | 1.29 | 1.32 |
| Kap Schelanija | 76°57′ | 68°34′ | 1.24 | 1.28 |
| Matotschkin Schar | 73°16′ | 56°24′ | 1.23 | 1.27 |
| Kap Schmidt | 68°55′ | 179°25′ | 1.22 | 1.20 |
| Jakutsk | 62°01′ | 129°43′ | 1.16 | 1.13 |
| Slutzk | 59°41′ | 30°29′ | 1.14 | 1.14 |
| Kursk | 51°45′ | 36°12′ | 1.12 | 1.10 |
| Aschchabad | 37°57′ | 58°23′ | 1.05 | 1.02 |

Die Messungen der Intensität der Sonnenradiation, die systematisch im Laufe des ganzen Jahres an verschiedenen Orten bei einer und derselben Sonnenhöhe durchgeführt werden, haben es ermöglicht einen genügend scharf hervorgehobenen Jahresverlauf des Radiationswertes mit einem Minimum in den Sommer- und einem Maximum in den Wintermonaten festzustellen, was wie durch die Veränderung des Wasserdampfinhaltes der Atmosphäre, so auch durch ihre Verstaubung im Jahresverlauf erklärt werden kann. Der gleiche Verlauf wurde wie in der Union, so auch im Auslande für Orte, wo aktinometrische Messungen ausgeführt wurden, erhalten. Alle die Orte, für die ein solcher Jahresverlauf der Intensität der Sonnenradiation erhalten worden ist, sind verhältnismässig in nicht hohen geographischen Breiten gelegen.

Das Vorhandensein von schon ziemlich lange andauernden systematischen aktinometrischen Beobachtungen in unserem Sektor der Arktis, gestattete den Jahresverlauf des Wertes der Intensität der Radiation (Durchsichtigkeit der Atmosphäre) auch in diesem Gebiet der Erdkugel zu erforschen.

Um mittels der oben angezeigten Methode die Änderungen der Durchsichtigkeit der Atmosphäre im Jahresverlauf nachzuspüren, ist es notwendig die Messungen der Intensität der Sonnenradiation bei einer und derselben Höhe der Sonne während des ganzen Jahres durchzuführen. Diese Bedingung kann für Orte, die nordlicher des Polarkreises gelegen sind, nicht eingehalten werden. Deswegen ist für die Messungen eine solche Höhe der Sonne gewählt worden, die für die allergrösste Anzahl der Monate im Jahre anzutreffen ist, die aber dennoch nicht so klein wäre, um die Genauigkeit der zu vollführenden Messungen zu beeinträchtigen. Als eine dem entsprechende Höhe ist die Höhe der Sonne von 14°3′, was einer «Masse» 4 entspricht, gewählt worden.

Für die Bucht Tichaja ist ein Beobachtungsmaterial von 5 Jahren von 1933 bis 1938, 1937 ausgeschlossen, und für die Insel Ujedinenija das Material für die Jahre 1935 und 1938 durchgearbeitet worden. Die für diese zwei Orte durchgeführte Durcharbeitung des Materials hat gezeigt, dass für sie ein Jahresverlauf der Durchsichtigkeit der Atmosphäre beobachtet wird, der invers zu dem ist, der uns von früher her, durch die Beobachtungen an südlicher gelegenen Orten, bekannt war.

In der Tab. 2 sind die Werte der Intensität der Sonnenradiation nach Monaten für die Bucht Tichaja und die Insel Ujedinenija, wie auch zum Vergleich für vier südlicher gelegene Orte unserer Union, angegeben.

Vergleicht man die Werte der Intensität der Sonnenradiation für Juli mit den Werten für März, so ergibt sich, dass für die Bucht Tichaja und die Insel Ujedinenija die Radiation im Juli grösser als im März beziehender

Tabelle 2

Mittlere monatliche Intensität der Sonnenradiation für die «Masse» 4, auf die mittlere Entfernung zwischen der Erde und der Sonne bezogen, in Kalorien

| Ort | I | II | III | IV | V | VI | VII | VIII | IX | X | XI | XII |
|---|---|---|---|---|---|---|---|---|---|---|---|---|
| Bucht Tichaja . . | | | 0.93 | 0.96 | 0.99 | 1.02 | 1.02 | 0.99 | | | | |
| Insel Ujedinenija . | | | 0.90 | 0.97 | 1.01 | 1.03 | 1.04 | 1.01 | | | | |
| Jakutsk . . . . . | | 1.00 | 1.02 | 0.99 | 0.89 | 0.86 | 0.85 | 0.87 | 0.99 | 1.02 | | |
| Slutzk . . . . . | | 1.00 | 0.99 | 0.91 | 0.86 | 0.85 | 0·79 | 0.82 | 0.91 | 0.97 | 1.00 | |
| Kursk . . . . . | 1.04 | 1.00 | 0.99 | 0.90 | 0.81 | 0.80 | 0.74 | 0.78 | 0.81 | 0.90 | 1.03 | 1.07 |
| Aschchabad . . . | 0.89 | 0.88 | 0.81 | 0.74 | 0.74 | 0.72 | 0.65 | 0.69 | 0.77 | 0.81 | 0.84 | 0.84 |

Weise um 9% und um 14% ist; für alle anderen Orte der Tabelle ist sie dagegen im Juli kleiner als im März, wobei diese Verkleinerung in der Reihenfolge der Lage der Orte in der Tabelle 2 beziehender Weise um 16, 20, 25 und 20% sein wird.

Aus der Abb. 1 kann ersehen werden, um wieviel für die Orte, die in der Arktis gelegen sind, der Jahresverlauf der Intensität der Sonnenradiation von einem gleichen Verlauf für südlichere Orte abweicht—er ist invers.

Der beobachtete Jahresverlauf der Intensität der Sonnenradiation wird im Grunde durch die Jahreskurve des Wasserdampfinhalts der Atmosphäre erklärt. Für die Bucht Tichaja und die Insel Ujedinenija ist aber der Jahresverlauf der absoluten Feuchtigkeit dennoch bedeutend, so verändert er sich für die Bucht Tichaja von 0.5 bis 7 mm; der erwartete Einfluss dieses Verlaufes auf die Intensität der Radiation wird aber nicht verspürt.

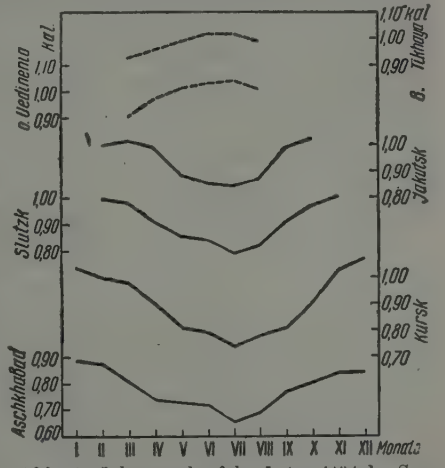


Abb. 1. Jahresverlauf der Intensität der Sonnenradiation für «Masse» 4, auf die mittlere Entfernung zwischen Erde und Sonne bezogen.

Der inverse Gang der Durchsichtigkeit der Atmosphäre in der Arktis wird scheinbar durch die Vergrösserung des Kondensationsdunstes der Atmosphäre in der kalten Jahreszeit zu erklären sein. Derselbe Verlauf der Durchsichtigkeit wie in der Arktis müsste auch in Jakutsk beobachtet werden, wo zur kalten Jahreszeit die Temperaturen niedrig sind; aus der Abbildung ist aber zu ersehen, dass in Jakutsk die Kurve des Jahresverlaufes von derselben Gestalt ist, wie für südlichere Orte. Vielleicht kann das durch kleineren Kondensationsdunst in Jakutien erklärt werden, weil dort die absolute Feuchtigkeit zur kalten Jahreszeit mehrfach kleiner ist, als in den Gegenden der Franz-Joseph Inseln und der Insel Ujedinenija.

Irgendwo in den geographischen Breiten zwischen Slutzk und der Insel Ujedinenija muss sich der Jahresverlauf der Intensität der Sonnenradiation durch eine Kurve, die nahe einer Geraden ist, darstellen lassen. Weit voneinander entfernte Lage der aktinometrischen Stationen am Norden und die geringe Zahl der Beobachtungsjahre der zur Zeit beschriebenen Stationen

gestattet heute leider nicht diese Frage zu beantworten; die Durchsicht des zur Zeit verfügbaren Materials zeigt aber, dass ein solcher gradliniger Jahresverlauf der Sonnenradiation in der Gegend der Breiten 71—73° zu liegen kommen muss.

Weitere Durchführung der systematischen Beobachtungen über die Intensität der Sonnenradiation, wie in der Arktis selbst, so auch in den nördlichen Gebieten unserer Union, wird diese interessante Frage genauer in den Einzelheiten zu erforschen gestatten.

Institut für Erforschung der Arktis.
Leningrad.

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*GEOPHYSIK*

# ÜBER DIE ENTSTEHUNGSFORM DES WASSERDAMPFKONDENSATS AN METALLFLÄCHEN BEI NIEDRIGEN TEMPERATUREN (UNTER 0°C)

By W. A. BASIKAILO

*(Vorgelegt von S. I. Wawilow, Mitglied der Akademie, den 24. II. 1941)*

In einer 1935 veröffentlichten Arbeit, die der Untersuchung von Psychrometern bei Temperaturen unter dem Wassergefrierpunkt (bis zu —20° C) gewidmet war, entnahmen Owbery und Griffiths (1) den Wert des partiellen Wasserdampfdruckes, der von ihnen mit Hilfe eines Kondensationsgygrometers bestimmt wurde, unmittelbar aus den Sättigungstabellen für einen gegebenen «Taupunkt» nicht über Wasser, sondern über Eis. Sie waren nämlich der Ansicht, dass die Kondensation des Wasserdampfes an stark abgekühlten Gegenständen über die flüssige Phase hinweg verläuft (Resublimation) (2). Aber bei der gleichen Temperatur der Wasser- und Eisoberflächen ist die maximale Wasserdampfspannung über dem Eis geringer, als über dem Wasser, und so stimmen die Werte, die sie für die Luftfeuchtigkeit nach dem «Taupunkt» und nach dem psychrometrischen Verfahren erhalten hatten, nicht überein; die letzteren waren etwa 10—15% höher, als die mit dem Kondensationsgygrometer ermittelten.

Ein derartiges Auseinandergehen der Ergebnisse nach den beiden üblichen Verfahren liess Owbery und Griffiths auf die Ungenauigkeit der Psychrometertafeln schliessen, auf denen sich der hydrometeorologische Dienst in UdSSR und Deutschland basiert. Es wurden von ihnen berichtigte Tafeln für Assmann'sche Aspiration aufgestellt, die geringere Luftfeuchtigkeit für mit trockenen und nassem Thermometer erhaltene Ablesungen ergaben. Diese Versuche waren in dem National Physical Laboratory Englands durchgeführt worden. Nun haben wir uns zur Aufgabe gestellt, ihre Ergebnisse zu prüfen. Als wir den Bildungsvorgang des Wasserdampfkondensats auf der polierten Oberfläche des Gerätes beim Taupunkt studierten, konnten wir uns davon überzeugen, dass an der Oberfläche eines Metallkörpers, wenn er bis zum Taupunkt herabgekühlt war, eine wirkliche Kondensation sogar dann stattfand, wenn die Temperatur tief unter dem Nullpunkt lag. Es wird aber allgemein angenommen, dass unter dem Triple-Punkt keine Kondensation, sondern eine Resublimation erfolgt. Die Versuche, die es ermöglichten die Kondensationserscheinung zu beobachten, liefen auf folgendes hinaus. Es wurden runde Metallplatten genommen, die sehr dünn—bis zu 0.1 mm—waren und einen Durchmesser von 12 bis 20 mm hatten; an einer Seite wurde die Oberfläche dieser Platten spiegelglatt poliert, an die Rückseite wurde ein Teil des Thermoelementes angelötet, dessen zweiter Teil in ein Bad mit tauendem Eise gelegt wurde. Diese Platten wurden sodann an einen dünnwandigen Ebonitzylinder angehaftet und bildeten einen Napf, der mehrere Tropfen einer leicht verdunstenden Flüssigkeit (Äther, Benzin) enthielt. Durch Verdunsten dieser Flüssigkeit wurde die Scheibenoberfläche gleichmässig abgekühlt. Ein in den Kreis eingeschaltetes Galvanometer ermöglichte es die

Änderung der Temperatur der Scheibe (des Spiegels) mit einer Genauigkeit von 0.1° C zu beobachten.

Diese Einrichtung gab uns die Möglichkeit, auf eine einfache und bequeme Weise eine langsame Abkühlung der Scheibe des Spiegels bis zu sehr niedrigen Temperaturen hinab zu erzielen. Bei der Abkühlung der Scheibe bis zum Taupunkt, die in unseren Versuchen einer Temperatur von —36° C entsprach, bildeten sich an ihrer Oberfläche Tropfen der durch Verdichtung von Wasserdampf entstandenen Flüssigkeit. Diese Kondensationsprodukte erschienen und verschwanden wieder, wenn die Temperatur des Spiegels um nur 0.1—0.2° C von derjenigen abwich, bei welcher die Kondensation noch stattfand. Dieses wurde beobachtet, wenn die flüssige Phase bis zum völligen Verschwinden des Kondensats dauerte. Bei niedrigen Temperaturen (unter—20° C) fand bisweilen eine spontane Bildung der festen Phase aus der flüssigen statt, doch nur an sehr beschränkten Teilen der Spiegeloberfläche, hauptsächlich an den Riss- und Kerbestellen. Hierbei befreite sich die Spiegelfläche um den Eiskristallen vom Eis in einem Kreise, dessen Halbmesser von 0.5 bis 0.8 mm betrug. Bei wiederholten Abkühlungen des Spiegels bis zum Taupunkt liess sich in der Nähe von den spontan entstandenen Eiskristallen keine sichtbare Kondensation bemerken. Die Verdunstung des Kondensats, der sich an der Scheibe in der Gestalt von einzelnen Kristallen gebildet hatte, wurde nur dann beobachtet, wenn die Temperatur der Scheibe um 1—2.5° C höher war, als die Temperatur, welche noch eine Kondensation ermöglichte. Beim Berühren des Kondensats mit leichten Gegenständen, wie Baumwolle- oder Papierfasern, dünnes Draht usw., ging die ursprüngliche flüssige Phase des Kondensats stets in die feste Phase über, d. h. ins Eis, aber dies geschah lediglich an den Berührungsstellen. Baumwollenfasern froren an die Spiegeloberfläche fest an. Das Gefrieren trat nicht sogleich ein, sondern erst nach 3—6 und bisweilen auch mehr Sekunden, in Abhängigkeit davon, wie tief der Taupunkt lag und wie dick das Kondensat an der Spiegeloberfläche war. In der Nähe der berührten Stellen, d. h. neben denjenigen Kondensatzonen auf dem Spiegel, wo wir durch Einbringen des «Zünders» die Kondensation des Wassers hervorgerufen hatten, war in der Entfernung von 0.5—0.8 mm von den Eiskristallen keine sichtbare Kondensation zu bemerken. Kleine Eiskörner, die auf den Spiegel aufgebracht wurden, nachdem letzterer bis zum Taupunkt abgekühlt war, benahmen sich hinsichtlich des Kondensats in derselben Weise, wie auch diejenigen, die spontan entstanden oder von uns durch Berühren künstlich bewirkt waren. Die Verdunstung der festen Kondensatform, die von uns künstlich durch leichtes Berühren der ursprünglichen Kondensatform ins Leben gerufen wurde, fand immer bei einer im Vergleich zum Taupunkt höheren Temperatur statt, und zwar um 1—2.5° C über derselben, je nach dem partiellen Wasserdampfdruck in dem Zeitpunkt der Beobachtung. Dieser Umstand, d. h. die höhere Verdunstungstemperatur der offenbar festen Kondensatform, weist darauf hin, dass sich die ursprüngliche Form in der flüssigen und nicht in der festen Gestalt bildet, wie man es bis nun anzunehmen pflegte. In unseren Versuchen reichte der Taupunkt bis—35.6° C hinab, und in allen Fällen, wo die Luftfeuchtigkeit unter natürlichen Verhältnissen an der Meteorologischen Warte in Slutzk beobachtet wurde, nahm man stets ein Kondensat wahr, das als dünner Film während einiger Minuten in flüssiger Form leben konnte.

Meteorologische Warte.
Slutzk.

Eingegangen
den 9. IV. 1941.

ZITIERTE LITERATUR

[1] Owbery and Griffiths, Proc. Phys. Soc., 47, 216 (1935). [2] Н. Н. Зубов, Морские воды и льды, 60 (1938).

*Übersetzt von K. Klein.*

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*CHEMIE*

# ZUR KONSTITUTION DES OXYPACHYCARPINS

Von A. P. ORECHOFF Mitglied der Akademie, M. I. KABATCHNIK und T. J. KEFELY

Pachycarpin wurde erstmals 1933 aus den Blättern und Stengeln der in Mittelasien einheimischen Pflanze, *Sophora pachycarpa*, isoliert. Die von A. P. Orechoff, M. S. Rabinowitch und R. A. Konowalowa (1) ausgeführte Untersuchung des Alkaloids ergab, dass es sich um den optischen Antipoden des Sparteins handelt.

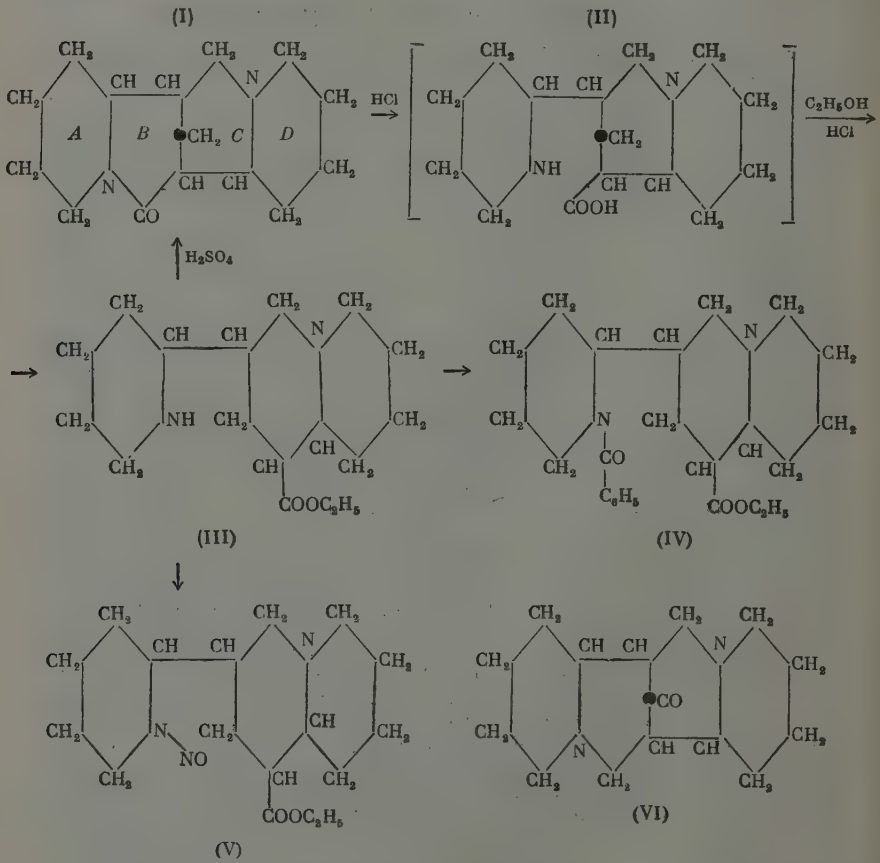
Durch Oxydation des Pachycarpins mit Kaliumferricyanid nach der etwas abgeänderten Schöpf'schen Methode (2) wurde Oxypachycarpin, das dem Ahrens'schen Oxyspartein (3) stereoisomer ist, erhalten. Für Oxyspartein wird gegenwärtig die Struktur (I) angenommen, da dieses ein Oxoderivat des Sparteins ist und die üblichen Reaktionen für die Carbonylgruppe nicht zeigt, was auf dessen Laktamnatur (4) hinweist. Die Stellung der Carbonylgruppe im Ring (*B*) wurde endgültig durch die Synthese des Oxysparteins nachgewiesen. Diese Synthese nimmt ihren Ausgang von Pyridyl-α-Essigsäure (5), wobei je zwei ihre Moleküle die Pyridinringe (*A*) und (*D*) des Oxysparteins liefern. Somit konnte die Carbonylgruppe sich nur in Ring (*B*) oder, was dasselbe ist, in Ring (*C*) befinden. Diese Tatsache, sowie das Ausbleiben der Carbonylreaktionen gestattete Clemo die Struktur (I) anzunehmen.

Es fehlte jedoch für die Laktamnatur des Oxysparteins ein direkter Beweis, umsomehr, als die Synthese von Clemo, die über eine Reihe nicht isolierter Zwischenprodukte verläuft, die Möglichkeit der Struktur (VI) für Oxyspartein offen lässt. Eine Substanz von dieser Konstitution könnte bei der Synthese von Clemo entstehen, wenn beide Moleküle des α-Pyridylessigesters in der Pyridon-Methidform reagieren würden (vgl. die Synthese von Clemo).

In vorliegender Arbeit stellten wir uns das Ziel, die Anwesenheit eines Laktamringes in dem Molekül des Oxypachycarpins und folglich auch in dem des Oxysparteins experimentell nachzuweisen.

Laktamringe in Alkaloiden, die dem Spartein verwandt sind, besitzen sehr verschiedene Stabilität. So ist das Laktam des Alkaloides Aphyllidin sehr beständig gegenüber Alkali, wird jedoch leicht von alkoholischer Salzsäure unter Bildung des Äthylesters der Aphyllidinsäure (6) gesprengt. Demgegenüber wird das Alkaloid Matrin leicht von Alkali gesprengt. Was das Lupanin betrifft, so ist dessen Laktamring ausserordentlich stabil, und es gelang erst unlängst E. Hoffman, F. W. Hohlschneider und K. Winterfeld (7) das Molekül des Lupanins durch Erwärmen mit konzentrierter Salzsäure im zugeschmolzenen Rohr auf 150° im Laufe von 72 Stunden zu sprengen.

Oxypachycarpin erwies sich als ausserordentlich beständig sowohl gegenüber der Einwirkung von Säuren, als auch von Alkalien. Erst beim Erwärmen mit konzentrierter Salzsäure im zugeschmolzenen Rohr auf 180° während 15 Stunden trat die gewünschte Sprengung ein. Zur Abtrennung der entstandenen Aminosäure (II) wurde diese in ihren Äthylester (III) verwandelt, der mit einer Ausbeute von 65%, bezogen auf das umgesetzte Oxypachycarpin, erhalten wurde. Der Äthylester (III) enthält eine Äthoxylgruppe und wird beim Verseifen mit 50%-iger Schwefelsäure in Oxypachycarpin zurückverwandelt. Hiermit ist das Vorhandensein einer Carboxylgruppe in dem Spaltungsprodukt bewiesen. Das Vorhandensein einer sekundären Aminogruppe wurde durch die Entstehung eines Benzoylderivates (IV), sowie eines Nitrosamins (V) bewiesen. Somit ist das Vorhandensein eines Laktamringes im Oxypachycarpin und damit die Struktur (I) für dieses Alkaloid bewiesen.




Sprengung des Oxypachycarpins; Äthylester der Pachycarpinsäure. 1 Teil des Oxypachycarpins wurde mit 5 Teilen konzentrierter Salzsäure im zugeschmolzenen Rohr 15 Stunden auf 180° erwärmt. Die erhaltene Lösung wurde mit 40%-iger Natronlauge stark

alkalisch gemacht (Lackmus). Das unveränderte Oxypachycarpin wurde mit Benzol extrahiert (es wurden 20—22% zurückerhalten), die alkalische Lösung filtriert, mit Salzsäure angesäuert und zur Trockne verdampft. Der trockene Rückstand wurde 10 Stunden bei 100° im Vakuum über $P_2O_5$ getrocknet und mit siedendem absolutem Alkohol ausgezogen. Der alkoholische Auszug wurde im Vakuum zur Trockne verdampft, der Rückstand mit 30 Teilen absoluten Alkohols übergossen und mit trocknem Chlorwasserstoff gesättigt. Die Lösung wurde eine Stunde lang auf dem Wasserbad unter Rückfluss erwärmt, wonach der Alkohol verdampft und der Rückstand in Wasser gelöst wurde. Die Lösung wurde mit 40%-iger Natronlauge stark alkalisch gemacht und mit Äther extrahiert. Der Auszug wurde über $Na_2SO_4$ getrocknet, der Äther verjagt und der Rückstand im Vakuum destilliert: Sp. 162—166°/2 mm. Ausbeute 65%. Der Äthylester der Pachycarpinsäure ist eine farblose, ölige Flüssigkeit, die sich beim längeren Aufbewahren blassgelb färbt.

0.1259 g Subst. gaben 0.1158 g $H_2O$ und 0.3218 g $CO_2$
7.363 mg bzw. 6.455 mg Subst. gaben 0.604 cm³ bzw. 0.532 cm³ $N_2$ (765 mm, 21°)
0.1794 g Subst. gaben 0.2040 g AgJ

$C_{17}H_{30}O_2N_2$ ber. C — 69.38%; H — 10.20%; N — 9.52%; $OC_2H_5$ — 15.30%
gef. C — 69.67%; H — 10.32%; N — 9.59%; 9.63%; $OC_2H_5$ — 15.01%

Drehungsbestimmung

0.2756 g Subst. in 4.6904 g Äthylalkohol ($l = 0.5$ dm)

$\alpha_D^{14} = -0.30°$; $[\alpha]_D^{14} = -12.2°$.

Bestimmung der molekulären Refraktion

$n_D^{20} = 1.5124$; $d_4^{20} = 1.062$; $C_{17}H_{30}O_2N_2$: *MR* ber. 83.29, gef. 83.24.

Ringschluss des Äthylesters der Pachycarpinsäure. Ein Teil des Äthylesters der Pachycarpinsäure wurde mit 10 Teilen 50%-iger Schwefelsäure bis zum Siedebeginn im Laufe von 3.5—4 Stunden erwärmt. Die abgekühlte Lösung wurde mit 40 Teilen kalten Wassers verdünnt und mit 20%-iger Natronlauge stark alkalisch gemacht. Die Base wurde mit Äther extrahiert, der Auszug über Natriumsulphat getrocknet und der Äther verjagt. Der Rückstand kristallisierte allmählich und wurde aus leichtem Petroläther umkristallisiert. Die erhaltenen Kristalle stellen nach ihrer Form, sowie nach Schmelzpunkt und Mischschmelzpunkt Oxypachycarpin dar.

Benzoylierung des Äthylesters der Pachycarpinsäure. Zu der Lösung von 1 g des Äthylesters und 5 cm³ trocknen Toluol wurde eine Lösung von 1 g Benzoylchlorid in 5 cm³ Toluol hinzugefügt und die Mischung 4 Stunden erwärmt. Nach dem Erkalten, erstarrte das Reaktionsgemisch kristallin. Es wurde Wasser hinzugegeben, die wässerige Schicht vom Toluol getrennt und mit Äther von Benzoylchlorid und Benzoesäure durch Extraktion befreit. Die wässerige Schicht wurde nun mit 20%-iger Natronlauge stark alkalisch gemacht und mit Äther extrahiert. Der getrocknete Ätherauszug hinterliess nach dem Verjagen des Lösungsmittels Kristalle, die nach Umkristallisieren aus Petroläther als lange farblose glänzende Nadeln vom Schmelzpunkt 121—122° erhalten wurden. Ausbeute 70% d. Th.

0.0961 g Subst. gaben 0.0748 g $H_2O$ und 0.2556 g $CO_2$
2.750 mg bzw. 3.770 mg Subst. gaben 0.169 cm³ (748 mm, 18°) bzw. 0.237 cm³ (748 mm, 17.5°) $N_2$

$C_{24}H_{34}O_3N_2$: ber. C — 72.36%; H — 8.54%; N — 7.04%
gef. C — 72.54%; H — 8.65%; N — 7.10 bzw. 7.27%

Nitrosierung des Äthylesters der Pachycarpinsäure. 1 g des Esters wurde in 10—15 $cm^3$ trocknen Äthers gelös. und die Lösung unter Eiskühlung mit Stickstoffoxyden gesättigt (2.5 Stunden) Der anfangs entstandene weisse Niederschlag ging gegen Schluss des Versuches wieder in Lösung. Die Lösung wurde mit gesättigter Sodalösung alkalisch gemacht und mit Äther extrahiert. Der Auszug wurde mit Natriumsulfat getrocknet und der nach dem Verjagen des Äthers verbleibende Rückstand im Vakuum (1 mm) bei 179—184° überdestilliert. Das Destillat kristallisierte allmählich. Ausbeute 52% d. Th. Nach zweimaligem Umkristallisieren aus trocknem Petroläther wurde die Substanz in Gestalt von dünnen farblosen langen Nadeln vom Schmelzpunkt 86—88° erhalten. Die Schmelze weist eine grüne Farbe auf. Das Nitrosamin zeigt die Liebermannsche Reaktion.

10.39 mg Subst. gaben 23.94 mg $CO_2$ und 8.09 mg $H_2O$
4.184 mg bzw. 2.777 mg Subst. gaben 0.465 bzw. 0.308 $cm^3$ $N_2$ (753 mm, 21° bzw. 22.5°)
$C_{17}H_{29}O_3N_3$: ber. C — 63.15%; H — 8.98%; N — 13.0%
gef. C — 62.84%; H — 8.71%; N — 12.78 bzw. 12.69%

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*CHEMIE*

# DIE ROLLE DER ÜBERSPANNUNG UND DES OHMWIDERSTANDES BEI DER PASSIVIERUNG VON BLEI

Von B. N. KABANOV

*(Vorgelegt von A. N. Frumkin, Mitglied der Akademie, den 31. I. 1941)*

Die Passivierung von Metallen durch Bildung an ihrer Oberfläche einer unlöslichen Salzschicht lässt sich bequem an dem Beispiele Blei+Schwefelsäure erforschen. W. Müller (1) hat eine Reihe von Untersuchungen auf diesem Gebiete veröffentlicht. Seine quantitative Theorie der Passivierung hat bisher nur wenig Widerspruch (2) gefunden; sie beruht auf der Annahme, dass der Anstieg des Elektrodenpotentials im Moment der Passivierung bei anodischer Polarisation ausschlieslich durch das Anwachsen des ohmschen Widerstands in den Poren bedingt ist, welche die passivierende Salzschicht durchsetzen.

Um diese Annahme zu beweisen, hat W. Müller den Widerstand der passiven Elektrode unmittelbar mittels Wechselstrom gemessen. Wir wollen diese Messung, welche W. Müller's Theorie des Passivzustandes begründet, hier näher betrachten. W. Müller hat beobachtet, dass der Widerstand einer passiven Elektrode einige Tausende Ohm pro 1 cm² ihrer Oberfläche beträgt*. Bei diesen Messungen wurde aber die Verschiebung der Phasen des Wechselstroms nicht in Betracht gezogen; der ermittelte Widerstand ist ein effektiver Widerstand und enthält ausser dem ohmschen auch den scheinbaren Widerstand, welcher durch die Beladung der elektrischen Doppelschicht und durch die Konzentration, sowie die chemische Polarisation der Elektrode hervorgerufen wird. Dieser scheinbare Widerstand addiert sich zum ohmschen, weil der Polarisations- und der Kapazitätswiderstand der Trennungsfläche zwischen Metall und Lösung in Serie mit dem ohmschen Widerstand geschaltet sind. Mit dem Anwachsen der Frequenz des bei der Widerstandsmessung verwendeten Wechselstroms vermindert sich der Teil des scheinbaren Widerstands, welcher durch die Kapazität der Doppelschicht bedingt ist; derjenige Teil, welchen die Konzentrationspolarisation bedingt, steigt dagegen an; der Teil, welcher durch die Überspannung bei der Entstehung oder Entladung von Bleiionen bedingt ist, hängt nicht von der Frequenz ab, ebenso wie der rein ohmsche Widerstand des Elektrolyten. Jedenfalls ist bei beliebigen Frequenzen der ohmsche Widerstand des Elektrolyten in den Poren der Blei-

* W. Müller und W. Machu (1) geben in ihrer jüngsten Abhandlung an, dass der Widerstand der passiven Bleielektrode 16000 Ω erreicht. Ihre Berechnung beruht auf Messungen mittels Gleichstrom. Vgl. auch die Untersuchung von Porfirow (3), der bei Anwendung von Wechselstrom einen Widerstand von 60—90 Ω pro 1 cm² beobachtete.

sulfatschicht geringer, als der Gesamtwiderstand, welcher auf gewöhnliche Weise mittels Wechselstrom gemessen wird.

Wir haben den Widerstand der mit einem passivierenden Salzfilm bedeckten Bleielektrode mit Berücksichtigung der Phasenverschiebung bestimmt. Im Frequenzbereich von 500 bis 6000 Hertz haben wir die Kapazität und den Widerstand einer glatten Bleielektrode* im Zustande der maximalen Passivität gemessen, d. i. bei einem Potential von ungefähr +1.3 V gegen die Wasserstoffelektrode. Als Lösung wurde dabei 8 *N* Schwefelsäure verwendet. Die Ergebnisse sind in Tab. 1 angegeben.

Tabelle 1

| Frequenz, Hz | Ohmsche Komponente | Effektiver Widerstand | Winkel der Phasenverschiebung |
|---|---|---|---|
| 500 | 142 | 287 | 60° |
| 3 000 | 28.5 | 60 | 61° |
| 6 000 | 19.2 | 36.6 | 58° |

Tabelle 2

| Frequenz, Hz | Ohmsche Komponente | Effektiver Widerstand | Winkel der Phasenverschiebung |
|---|---|---|---|
| 500 | 26 | 88 | 73° |
| 3 000 | 6.5 | 19 | 69.5° |
| 6 000 | 4.8 | 11.4 | 65° |

Ähnliche Resultate für den Anfangszustand der Passivierung (bei einem Potential von +0.45 V) sind in Tab. 2 gegeben.

Wie ersichtlich, vermindert sich die ohmsche Komponente mit dem Sinken der Frequenz; somit enthält sie ausser dem ohmschen Widerstand auch noch den scheinbaren. Der ohmsche Widerstand einer maximal passiven Elektrode ist geringer als 19 Ohm, und im Anfangsstadium der Passivierung geringer als 5 Ohm. Bei Stromdichten von weniger als 0.01 A/cm² ist also der ohmsche Potentialabfall in den Poren der passivierenden Schicht verschwindend klein im Vergleich mit der Polarisation der Elektrode im passiven Zustande**. Aber auch bei höheren Stromdichten wird das durch Messung ermittelte Elektrodenpotential nicht durch seinen ohmschen Widerstand bestimmt. Thermodynamische Erwägungen ergeben, dass bei anodischer Polarisation einer passiven Elektrode keine Ionen mit höherer Valenz entstehen können [entgegen den Behauptungen von W. Müller (⁴)], falls das Elektrodenpotential geringer ist, als das reversible Potential des Übergangs in ein Ion mit höherer Valenz; ferner, dass die Entstehung des letzteren durch den ohmschen Potentialabfall in der Lösung nicht erleichtert werden kann. Wäre der ohmsche Widerstand bedeutend, so würde das durch Messung ermittelte Gesamtpotential bei der Passivierung durch einen Strom von entsprechender Dichte das Gleichgewichtpotential des Bleisuperoxyds wesentlich überschreiten. Bei einem ohmschen Widerstand von 40 Ω z. B. könnte bei einer Stromdichte von 0.1 A/cm² die Entstehung von vierwertigem Blei erst bei einem gemessenen Potential beginnen, welcher um 4 V grösser ist als das positive Potential der Elektrode $PbO_2$—$PbSO_4$. In Wirklichkeit ist aber weder bei den oszillographischen Messungen von W. Müller, noch im Verlaufe unserer Untersuchung solches beobachtet worden. Somit haben die Versuche gezeigt, dass der ohmsche Widerstand an der Potentialänderung beim Passivieren keinen Anteil nimmt.

* Die Elektroden wurden vor dem Versuch anodisch passiviert und dann mittels des Kathodestroms reduziert; dadurch wurde ihre wahre Oberfläche ungefähr zweimal so gross, wie die scheinbare. Diese Grösse wurde durch Messung der Kapazität der Doppelschicht der reduzierten Elektrode bestimmt. Die ohmsche Komponente wurde an Hand eines Schemas der Hintereinanderschaltung berechnet.

** Unsere Versuche haben ergeben, dass für die Passivierung einer Bleielektrode und Wahrung des passiven Zustands (1.3 V) eine Stromstärke von der Grössenordnung $5 \times 10^{-6}$ A/cm² der scheinbaren Oberfläche ausreicht.

Wir wollen nun versuchen zu erklären, wodurch die Potentialverschiebung bei der Passivierung von Bleielektroden in Wirklichkeit bedingt wird. Betrachten wir die Bleipassivierung in 8 $N$ Schwefelsäure und nehmen (als erste Annäherung) an, dass $PbSO_4$ in die feste Phase unmittelbar an der Metalloberfläche übergeht. Ferner soll zunächst angenommen werden, dass die Stromdichte in den Poren diejenige des Diffusionsstroms für Schwefelsäure nicht überschreitet. In diesem Fall, der eingehend von Eucken (5) behandelt worden ist, kann man das Vorhandensein eines zweiten Elektrolyten ($PbSO_4$) ausser Acht lassen. Somit ändert sich in diesem Fall die Konzentration längs einer Pore bei der Entstehung der Konzentrationspolarisation geradlinig mit der Entfernung*.

Die Grösse des Diffusionsstroms wird durch die Formel

$$I_0 = \frac{3URTqC_0}{\delta} = \frac{3.68 \times 8.31 \times 298 \times q \times 4 \times 10^{-3}}{96500 \times 2.5 \times 10^{-4}} = 8.4q$$

ausgedrückt, wo $U = \frac{u}{F}$ — die absolute Beweglichkeit des Ions $SO''_4$ bedeutet, $q$—die gesamte Querschnittfläche der Poren, $C_0$—die Säurekonzentration in der Lösung ausserhalb der Poren (in Mol/cm³), $\delta$—die Länge der Poren, ungefähr gleich der Dicke der Bleisulfatschicht.

Bei einer Stromdichte von 42 A/cm² in den Poren (d. i. nur 2 Mal niedriger als der Grenzwert) ist die Konzentration an der Metalloberfläche $C_k$ nur 2 Mal geringer, als in der Mitte der Lösung.

Nach Eucken ist die Summe des ohmschen und des Diffusions-Potentialabfalles an der Elektrode unter diesen Bedingungen gleich

$$\varphi_1 - \varphi_2 = \frac{RT}{F} \ln \frac{C_0}{C_k} = 18 \text{ mV}.$$

Bei einer Stromdichte von 0.17 A/cm² ist die auf diesem Wege errechnete Summe des ohmschen und des Diffusions-Potentialabfalls gleich 0.1 mV. Andererseits liesse sich bei einer so hohen Stromdichte schon eine bedeutende Polarisation im Zusammenhang mit der Ionisationsüberspannung des Bleis erwarten. Deshalb haben wir die anodische Überspannung des Bleis bei hohen Stromdichten gemessen, und zwar unmittelbar nachdem sich ein stationäres Potential eingestellt hatte; zu dieser Zeit war die effektive Oberfläche der Elektrode, auf Kosten der Belegung durch eine Bleisulfatschicht, noch nicht merklich verringert. Bei der gleichen Stromdichte wurde die Überspannung des Wasserstoffs gemessen, und da die Überspannungskurve keine Biegung nach oben aufwies, kann man schliessen, dass während unserer Messungen die Konzentrationspolarisation und der ohmsche Potentialabfall in der Lösung nur noch ganz unbedeutende Werte erreicht hatten. Bei einer Stromdichte von 0.17 A/cm² (Abb. 1—die Kurve der anodischen

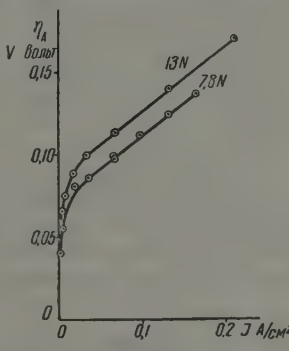


Abb. 1. Anodische Überspannung an Blei in 7.8 $N$ und 13 $N$ Schwefelsäurelösungen bei einer Bedeckung von 0.3 der Bleioberfläche mit Bleisulfat. $I$ ist die Stromdichte nach Umrechnung auf 1 cm² der Oberfläche einer glatten Elektrode.

* Um die hohen Zahlenwerte für den Widerstand in den Poren, welche sich auf Grund der Elektrodenpolarisation errechnen, erklären zu können, nimmt W. Müller ohne Begründung an, dass infolge der Konzentrationspolarisation der spezifische Widerstand der Lösung in den Poren um ein Vielfaches den Volumwiderstand übertrifft und dass er dem spezifischen Widerstand der gesättigten Salzlösung gleich ist.

Überspannung an Blei) beträgt die Überspannung bereits 140 mV, während der ohmsche und der Diffusions-Potentialabfall in den Poren verschwindend klein sind (0.1 mV)*. Somit liegt die Ursache der Potentialverschiebung bei einem passivierenden Strom von konstanter Stärke, wenigstens in den Anfangsstadien der Passivierung, in dem Anwachsen der Überspannung an Blei mit der Erhöhung der Stromdichte in den Poren.

Bei hohen Stromdichten ist die Messung der Überspannung schwerer auszuführen. Die Extrapolation des geradlinigen Teils der Überspannungskurve (Abb. 1) ergiebt z. B. für die Stromdichte von 1 A/cm² eine Überspannung von 470 mV bei $\varphi_1 - \varphi_2 = 3.6$ mV. Ist die Extrapolation hier zutreffend, so kann behauptet werden, dass im gesamten Bereiche der Potentialänderung bei der Passivierung (ungefähr 2 Volt) das Elektrodenpotential durch die Überspannung der Bleiionisierung bestimmt wird. Ehe die Stromdichte in den Poren noch eine Grösse erreichen kann, bei welcher der ohmsche Potentialabfall einen merklichen Wert erhält (etwa mehr als 10 mV), überschreitet das Elektrodenpotential die Grösse des Gleichgewichts-Potentials $PbO_2 - PbSO_4$ und beginnt die Bildung von Bleisuperoxyd.

Von diesem Standpunkt aus muss angenommen werden, das eine Potentialverschiebung, welche dem Übergang in den passiven Zustand entspricht, dann stattfindet, wenn die Stromdichte in den Poren eine gewisse Grösse erreicht hat, unabhängig von der Stärke des polarisierenden Stroms, d. h. der Gesamtquerschnitt der Poren hängt in diesem Moment von der Stromstärke ab. Ebenso muss auch der ohmsche Widerstand beim passiven Zustand der Elektrode von der Stärke des polarisierenden Stroms abhängen, und zwar wird er sich bei Verminderung der letzteren erhöhen.

Die von W. Müller und anderen gegebene Erklärung der Passivierungsursache für andere Systeme, als Blei-Schwefelsäure, muss ebenfalls von dem von uns entwickelten Standpunkte aus nachgeprüft werden. So nimmt, z. B. Kurtz (²) an, dass der ohmsche Widerstand die Änderung des Bleipotentials bei Passivierung des Bleis in Gegenwart von Chlorionen bedingt. In diesem Falle entsteht ebenfalls eine dünne Salzschicht (Dicke—etwa 1 μ); der Verlauf der Kurve für die anodische Überspannung an Blei in Gegenwart von Chlorionen darf nicht wesentlich von derjenigen abweichen, welche in Gegenwart von Schwefelsäureionen erhalten wird. Deshalb kann auch hier erwartet

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* Erdey-Gruz und Volmer nehmen an, dass die Überspannung an Metallen durch den ohmschen Widerstand der Lösung an den aktiven Stellen (Stufen) eines Kristalls der Metallelektrode bedingt ist. Es muss hervorgehoben werden, dass dieser Potentialabfall in der Lösung von dem vorhin betrachteten Abfall in den Poren der passivierenden Schicht wesentlich verschieden ist. Bei Messung mittels Wechselstrom muss man den Widerstand an den aktiven Stellen als der Kapazität des grössten Teils der Doppelschicht parallel angeschlossen betrachten, und den Widerstand in den Poren—als in Serienschaltung befindlich. Dementsprechend erhöht der Widerstand an der aktiven Stelle die Messungsergebnisse bei ihrer Berechnung gemäss dem Schema einer serienweisen Schaltung nicht wesentlich. Die Ursache der Passivität der übrigen Metalloberfläche liegt, nach Erdey-Gruz und Volmer, darin, dass für die Entstehung eines zweidimensionalen Keims an der Metalloberfläche eine bestimmte Aktivierungsenergie erforderlich ist. Demnach ist, gemäss dieser Theorie, die Entladung von Metallionen an inaktiven Stellen verzögert, und dies ist die erste Ursache der Überspannung an Metallen. Diese Theorie der Überspannung soll hier nicht näher betrachtet werden; wir vermerken bloss, dass sie unseren Daten nicht widerspricht: sie führt zu der Folgerung, dass der Potentialsprung an den aktiven Stellen zwar nur einen geringen Teil des totalen Potentialsprungs ausbildet und durch den grossen Potentialabfall in der Lösung zu dem experimentell gemessenen Wert der Überspannung ergänzt wird, dass aber der übrige Teil der Metalloberfläche an den Elektroden einen wahren Potentialsprung von bedeutender Grösse zeigt, gleich der gemessenen Überspannung. Somit kann beim Anwachsen der Stromdichte (und folglich auch beim Anstieg der Überspannung) der Potentialsprung an dem nichtaktiven Teil eines Metallkristalls einen solchen Zahlenwert erreichen, dass der Übergang des Metalls in Ionen von höherer Valenz thermodynamisch möglich wird.

werden, dass die Ursache für die Potentialverschiebung in der Überspannung an Blei liegt. Ebenso müssen auch W. Müllers Schlussfolgerungen in bezug auf die Passivierung von Eisen geprüft werden.

Dem Akademiker A. Frumkin sagen wir besten Dank für seine Ratschläge, welche der Durchführung dieser Untersuchung verholfen haben.

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*Übersetzt von A. Thal.*

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*PHYSICAL CHEMISTRY*

# INFLUENCE OF SORPTION UPON THE ELECTRICAL CONDUCTIVITY OF PULVERIZED SORBENTS

By N. A. FIGUROVSKY and M. M. FEINBERG

*(Communicated by A. N. Frumkin, Member of the Academy, 24. II. 1941)*

The electrical conductivity of a layer of pulverized active coal deposited on to a dielectric varies considerably in atmospheres containing vapours of various organic substances. Moreover, as a rule, the conductivity increases up to a certain limit value, depending on the nature and concentration of the vapour. If organic substances are eliminated from the air surrounding the coal, the conductivity decreases. Apparently, this phenomenon is related to sorption of the vapour to the surface of the coal grains.

It may have been expected that the sorption of the vapour of such substances as benzene, toluene, and other aromatic hydrocarbons, whose electrical resistance is several orders higher than that of coal, should decrease the conductivity of the coal layer. But the truth is that the opposite effect was observed by us in all cases, which shows that the electrical properties of the sorbed substance have no direct influence upon the conductivity of the pulverized sorbent.

Published data on the electrical resistance of sorbents with adsorbed vapour and gases are scarce and contradictory. Siebel [1] observed, that the resistance of preliminarily degased carbon filaments increases, due to sorption of air and ammonia. Kleine [2] obtained the same results for metallic filaments, and so did Bennewitz and Gunter [3]. Observations made by Suhrmann [4], Janitsky [5] and Germann [6] showed the contrary; according to these authors, sorption decreases the resistance. It must be noted, that in all the researches mentioned the observations were made on homogeneous materials—metallic or carbon filaments. We, on the other hand, have studied the electrical properties of pulverized coal, whose grains come into contact with each other only at individual points, instead of all over their surfaces. Thus, in the phenomena which we studied the main factor was the transitional resistance between grains.

Experimental Part*. We studied the electrical properties of pulverized coals deposited from suspension on to pumice, ebonite, porcelain and other dielectrics. These materials were in the form of cylinders 50 mm long and 6 mm in diameter. The ebonite cylinders had longitudinal milled grooves, into which the pulverized coal was pressed. Metal clamps were fastened to the ends of the cylinders, to guarantee stable contact with

* G. G. Edelmann took part in making the measurements.

the sorbent layer. Thus fashioned, this «element» was switched into the circuit of a Wheatstone bridge for direct current. The initial resistance of the elements varied from 1000 to 100000 ohms, depending on the thickness of the coal layer and the degree of dispersion of the powder.

The experiments were performed under conditions of static and dynamic sorption from air with vapours of benzene, toluene, aliphatic alcohols, ether, carbon tetrachloride and other substances in various concentrations.

Fig. 1 gives the curves of relative resistance decrease («depression») for three alcohols at a concentration of 10 mg per liter of air, in relation to the time of action upon the element. A comparison of these curves shows that the maximum depression value and the rate of resistance change increase with the molecular weight of the alcohol.

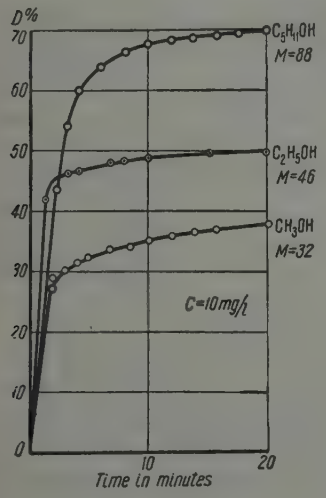


Fig. 1.

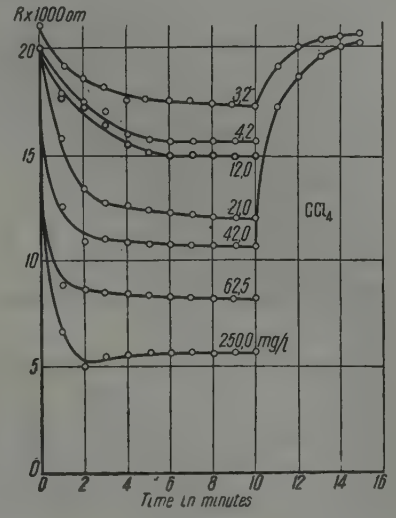


Fig. 2.

Fig. 2 shows the dynamics of resistance change for various concentrations of carbon tetrachloride. Increase of the concentration results in the rate of resistance decrease, the maximum depression value also increasing. At a concentration of 250 mg/l, the depression is as high as 74% i. e. the conductivity of the coal layer increases four times. If the element is transferred to pure air, the resistance rises again up to a value close to the initial one. This is probably due to desorption. However, if a number of successive experiments are performed with the same element, a gradual increase is observed both in the resistance of the element in air and in the rate and final value of resistance decrease in the vapours of organic substances.

The experiments were performed on active coals of various brands and degrees of dispersion. The best results were obtained with birch charcoal. As to the degree of dispersion of the powders, it was found that neither coarse, nor highly dispersed powders give good results. The most sensitive powders are those with particles from 50 to 100 μ, which can easily be isolated by elutriation.

Besides the elements just mentioned, we studied elements, prepared by depositing coal on porous glass, by filtering a coal suspension through a Schott filter, and also by baking pulverized glass in which coal par-

ticles were uniformly distributed. In the latter case the contacts were of silver and were fused into the porous glass.

The nature of the sorbed substance also influences the rate and the final value of the resistance decrease. Thus, the sensitivity of the elements to polar substances, such as organic haloid derivatives, is higher than to saturated hydrocarbons.

The number of experiments performed is as yet insufficient to explain the mechanism of this phenomenon fully. However, some preliminary statements can already be made.

The change in resistance of the coal layer is regularly related to the sorption and desorption of organic vapour from the air. The speed with which equilibrium is reached, shows that the reversible sorption, responsible for the resistance drop in the coal layer, takes place mainly on the surface and in the macropores of the coal.

One of the possible explanations of the mechanism of the resistance drop can be made by assuming that sorption increases the volume of the coal particles, thereby improving the contact and decreasing the transitional resistance between them. This assumption is confirmed by the fact that coal has been observed to swell upon sorption. Such observations have been made by Pavlov (7), Mc Bain and Porter (8), Bangham and Fakhoury (9), etc. In experiments with graphite and lamp black we observed that the resistance increases in the vapours of organic substances.

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*GEOCHEMISTRY*

# RADIOACTIVITY OF THE WATERS AND ROCKS OF MATZESTA

By M. S. MERKULOVA

*(Communicated by V. G. Chlopin, Member of the Academy, 24. I. 1941)*

Until now we had no sufficient radiochemical characteristics of the mineral waters of the Matzesta region. The knowledge of the radioactivity of the waters and rocks, however, is of a doubtless interest, both for a fuller knowledge of the mineral waters of the given region, and for the solution of some questions of their genesis.

The springs of the Old and New Matzesta, as well as of Agura, represent a thermal rather highly mineralized chloride water with a high content of free and combined hydrogen sulphide. According to the analyses made at the Chemical Laboratory of the Institute of Medical Hydrology, the hydrogen sulphide content of the given waters is on the average 0.150 g; the chlorine content, 4 to 7 g; the total mineralization, 6 to 12 g per liter. All the wells of Agura-Matzesta yield water of an essentially similar composition. The difference between the individual springs lies only in the total concentration of salts and in the hydrogen sulphide content. The waters of Agura-Matzesta are distinguished by a negligible content of sulphides and a high content of ammonium, iodine and boric acid. The $SO_4''$ concentration in individual wells varies within the range of 0.2 to 1.8 mg per liter, with the exception of the wells 5 and 6 of Old Matzesta, where a $SO_4''$ content of 8 to 10 mg per liter has been found. Near to the Agura-Matzesta waters by their chemical composition are the springs of the southern region, namely the Mendji springs. The slightly mineralized chloride waters of Aibga are essentially different from the Matzesta waters in chemical composition. These waters contain hydrogen sulphide in an amount of 4 to 5 mg per liter, $NH_4'$, $I'$ and $Br'$ being absent.

We have investigated for their Ra and $Msth_I$ content the waters of Old and New Matzesta, of Agura, of Bzugu well No. 30, as well as the springs of Mendji, Aibga, etc. The radiochemical analysis of the given waters was made by the usual method, by isolating Ra and its isotopes with barium salts. The volume of waters, from which Ra and its isotopes were isolated, was 1 to 4 liters. The measurements of Ra and $Msth_I$ were made on the compensation apparatus by means of an Lütz-Edelmann electrometer. The measurements of $Msth_I$ were made after six months, a year, and a part of the samples after 1 year 4 months, after sampling. The data obtained on the Ra content are given in Table 1.

Table 1

| Bore-hole or spring | $t°$ of water | Discharge 1000 liters per 24 hours | Ra content in grams per liter of water $\times 10^{11}$ | Content of emanation in eman * |
|---|---|---|---|---|
| **Old Matzesta** | | | | |
| Bore-hole No. 2 | 24.0 | 69.8 | 2.1 | 9.90 |
| » No. 3 | 26.0 | 240.9 | 2.2 | 6.00 |
| » No. 4 | 27.0 | 246.5 | 3.4 | 3.49 |
| » No. 5 | 22.3 | 115.7 | 2.1 | 7.76 |
| » No. 6 | 22.4 | 57.9 | 2.5 | 6.10 |
| » No. 7 | — | — | 3.4 | 1.24 data refer to bore-hole 6/7 |
| » No. 14 | 22.7 | 213.3 | 0.9 | 9.65 |
| **New Matzesta** | | | | |
| Bore-hole No. 9 | 31.4 | 360.0 | 4.7 | 2.0 |
| » No. 13 | — | — | 0.2 | 1.52 |
| Agura, bore-hole No. 6 | 28.8 | — | 8.5 | 8.40 |
| » No. 7 | 29.0 | 919 | 3.1 | — |
| » No. 9 | 30.2 | 2400 | 8.9 | 5.84 |
| Bzugu, bore-hole No. 30 | — | 960 | 1.3 | — |
| Mendji, bore-hole No. 9 | 21.0 | — | 4.6 | — |
| Spring No. 739 | 16.0 | — | 1.4 | — |
| Griffon No. 87 | — | — | $< 0.07$ | — |
| Aibga, spring No. 1 | 13.2 | — | 0.5 | — |
| » Mindolinsky spring No. 2 | 14.7 | — | $< 0.07$ | — |
| » spring No. 3 | 15.4 | — | $< 0.07$ | 7.30 |
| » » No. 8 | — | — | $< 0.07$ | — |

As it will be seen from Table 1, the waters of the Old and New Matzesta, as well as of Agura, which are characterized by a high mineralization and a negligible content of $SO_4''$ ions, contain measurable quantities of radium. The radium content in these waters varies from 0.9 to $8.9 \times 10^{-11}$ g of Ra per liter of water. The slightly mineralized waters of Aibga, differing in chemical composition, contain practically no radium. Their radium content in all the cases, with the exception of spring No. 1, is lower than the sensitivity of the method used.

Thus, the above data enable us to establish a certain relationship between the chemical composition of the waters and their radium content. Passing now to the occurrence of the products of the thorium series in the same waters, it is necessary to note that their content lies in all the cases beyond the limits of sensitivity of the method. The Th content in all the cases was found to be less than $5 \times 10^{-6}$g of thorium per liter of water.

In addition to the radiochemical analysis of the waters of Matzesta and the region, there were also made analyses of the rocks associated with the waters, for their radium and thorium content. The radiochemical analysis of the rocks was made by the method of Prof. I. E. Starik (1).

Table 2 shows the values of the content of Ra and Th in the principal rocks of the region.

* Data of A. P. Ratner. Unpublished report of the expedition of the Institute of Medical Hydrology, 1937.

Table 2

| Rock | Age of rock | Content in gs per 1 g of rock | | Coefficient of emanation | |
|---|---|---|---|---|---|
| | | Ra | Th | Ra | Th |
| Marl . . . . . | Lower Albian | $2.08 \times 10^{-12}$ | $0.5 \times 10^{-5}$ | 31.8% | 35% |
| Limestone . . . . . | Upper Cretaceous, Senonian | $0.69 \times 10^{-12}$ | $< 0.5 \times 10^{-5}$ | 25.3% | — |
| Limestone . . . . . | Upper Cretaceous, Turonian | $0.14 \times 10^{-12}$ | $< 0.5 \times 10^{-5}$ | — | — |
| Limestone compact . | Lower Cretaceous | $0.32 \times 10^{-12}$ | $< 0.5 \times 10^{-5}$ | — | — |
| Calcareous sandstone | Lower Cretaceous, Cenomanian | $< 0.10 \times 10^{-12}$ | $< 0.5 \times 10^{-5}$ | — | — |

From the above table it will be seen that the values of the Ra and Th content in the rocks approach the average values observed for sedimentary rocks. Out of the specimens studied, the most active are marls which contain $2.08 \times \times 10^{-10}$% Ra. In two specimens of rocks the coefficients of emanating of the radium and thorium emanation were determined and were found to be on the average 28.5% for the radium emanation, and 35% for the thorium emanation. Thus, we see that while the Ra content in the waters of Matzesta is rather high, neither the content of radium itself nor the content of radon is so high as to constitute an independent factor in medical hydrology.

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*CHEMICAL TECHNOLOGY*

# CATALYTIC AGENTS FOR THE PRODUCTION OF CARBON DISULPHIDE FROM THE ELEMENTS

By L. J. MARKOVSKY

*(Communicated by E. V. Britzke, Member of the Academy, 15. II. 1941)*

Reactions of carbon oxidation, whose mechanism apparently consists in a combination of topochemical and surface reactions ([1–5]), are subject to pronounced catalytic action of various oxides and salts. The velocity of these reactions is enhanced considerably by the presence of certain metallic oxides or ions (especially, those of alkali and alkaline earth metals) on the surface of the carbon ([6,7]). The mechanism of this catalysis consists most likely in the acceleration of the velocity of decomposition of very stable surface oxides which form on the surface of the carbon.

Studies of the catalytic reactions of carbon are very interesting from a theoretical standpoint, since they throw light upon the mechanism of various reactions of the latter.

The analogy between the chemical properties of oxygen and sulphur, and that between the structure of $CO_2$ and $CS_2$ suggested that catalysis may play some part in the reaction between carbon and sulphur during the production of carbon disulphide. This problem is interesting, besides, from a practical standpoint, since the formation of carbon disulphide from carbon and vapourized sulphur is a very sluggish reaction and in industry requires the most active kinds of carbonaceous materials (charcoal, peat coke). The use of catalyzers should increase the capacity of the apparatus.

The method of comparative study of the reactive capacity of carbonaceous materials, which had already been worked out ([8]) for reactions of carbon disulphide formation, made it possible to study the problem of selecting catalyzers for this process.

The salts (mostly sulphates and carbonates) of the following metals were studied: Na, K, Ca, Fe, Ni, Mn, Cu; besides these, $SiO_2$ was also investigated. Most of them, except K and Na, either have no effect on the velocity of the reaction between carbon (in the form of charcoal) and sulphur vapours, or even decrease it. On the other hand, K and Na salts, as the table below shows, increase the rate of reaction and also the yield of $CS_2$. Experiments with charcoal which had been preliminarily freed from ash by treatment with acids (ordinary charcoal contains 1—2% ash, about half of which is alkali) showed that the high chemical activity of charcoal was due in part to the presence of alkali components in its ash. Mineral admixtures were introduced into the charcoal by impregnating its preliminarily out-gased portions with solutions of the respective salts, after which the charcoal was dried and calcined.

These data show that the yield of carbon disulphide can be increased 20—30% by introducing 3—4% of alkali carbonates or sulphides into the charcoal. If the salt content is higher than this, the yield of $CS_2$ gradually drops, owing probably to a decrease in the free surface of carbon.

These data were checked at one of the carbon disulphide plants with the cheapest of the admixtures studied—sodium sulphate. The verification fully confirmed the data obtained in the laboratory, and showed that if 3% $Na_2SO_4$ is added to the charcoal with which the retorts are charged, the capacity of the latter in the production of carbon disulphide is increased not less than 30%. Since the addition of catalytic agents in no way changes the technology of the process and does not lead to extra expense, there can be no doubt as to the practical importance of these results. Laboratory experiments show that the action of the admixtures consists not only in activizing the surface of the charcoal, by oxidizing it partly during their reaction with it. Charcoal, which had been preliminarily calcined at the required temperature together with the admixture and then carefully washed free of the latter, displays no effect at all. Thus, it is necessary that the salts be present during the reaction.

Action of Admixtures of Salts of Alkali Metals upon Relative Yield of $CS_2$

| | Yield of $CS_2$ in % of charcoal |
|---|---|
| Charcoal (birch) | |
| Containing 1% ash . . . . . | 100 |
| Charcoal, free from ash . . . | 91.6 |
| » + 3% $Na_2CO_3$ . . . | 123 |
| » + 3% $K_2CO_3$ . . . . | 123 |
| » + 4.5% NaOH . . . | 128 |
| » + 4% $Na_2SO_4$ . . . | 123 |
| » + 4% $Na_2CO_3$ . . . | 131 |

The data obtained in this work give grounds to assume, that the formation of carbon disulphide passes through intermediate stages which consist in the formation and decomposition of fairly stable surface sulphides, analogous to those studied by Wibaut ([9]). The catalytic action of the admixtures studied is most probably due to their accelerating the decomposition of these surface sulphides; the mechanism of this process consists in the formation of topochemical compounds of carbon with alkali metals ([10], [11]), or with their fused salts, sulphides in this case, which weakens the valency bonds between the carbon atoms in the graphite lattice.

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*CRYSTALLOCHEMISTRY*

# THE CRYSTALLINE STRUCTURE OF $Zn(CN)_2$

By H. S. ZHDANOV

*(Communicated by A. N. Frumkin, Member of the Academy, 20. II. 1941)*

In connexion with an examination of the crystal structure of complex cyanides we have determined the structure of zinc cyanides. The experimental data, the spacing $d$ of the reflecting planes and the intensities of the X-ray lines, measured on a powder photograph (exposed on a tube with Mo-target and with Zr-filter) were taken from the work of Hanawalt, Rinn and Frevel (1). These data are given in column 2 of Table 1 and in column 4 of Table 2.

It is possible to index all the observed lines assuming a cubic lattice with the unit cube edge $a=5.89$ Å (Table 1).

Table 1

Indexing of the Powder Photograph of $Zn(CN)_2$

| Nos. | $d$ | $hkl$ | $a$ | Nos. | $d$ | $hkl$ | $a$ | Nos. | $d$ | $hkl$ | $a$ |
|---|---|---|---|---|---|---|---|---|---|---|---|
| 1 | 2 | 3 | 4 | 1 | 2 | 3 | 4 | 1 | 2 | 3 | 4 |
| 1 | 4.19 | 110 | 5.92 | 6 | 1.86 | 310 | 5.88 | 11 | 1.315 | 420 | 5.88 |
| 2 | 3.40 | 111 | 5.89 | 7 | 1.70 | 222 | 5.89 | 12 | 1.256 | 332 | 5.89 |
| 3 | 2.95 | 200 | 5.91 | 8 | 1.57 | 321 | 5.875 | 13 | 1.202 | 422 | 5.89 |
| 4 | 2.40 | 211 | 5.88 | 9 | 1.47 | 400 | 5.88 | 14 | 1.157 | 510 | 5.90 |
| 5 | 2.08 | 220 | 5.88 | 10 | 1.39 | 411 | 5.90 | $a_m = 5.89$ Å | | | |

The Bravais lattice is primitive. The indices of all the observed lines, except the second line, satisfy the condition $h+k+l=2n$. This fact shows that the chief scattering centres in the structure of $Zn(CN)_2$ form a body-centered lattice.

The number of the molecules $Z=2$ in the unit cube gives reasonable value of the density. Therefore we must place the following atoms: 2 Zn, 4 C and 4 N.

Examination of the space groups of the cubic system with a primitive lattice has shown that to realize the distribution of groups of 2 and 4 atoms is possible in the following 5 groups only: $T^1$, $T_h^2$, $T_d^1$, $O^2$ and $O_h^4$.

It is easy to exclude groups $T_h^2$, $O^2$ and $O_h^4$ which have two- and fourfold positions without parameters. Thus, only 2 space groups $T^1$ and $T_d^1$, in which

the examined special positions are identical, remain for the subsequent analysis. These positions are as follows:

$$1:(a)\,000;\quad 1:(b)\,\tfrac{1}{2}\tfrac{1}{2}\tfrac{1}{2}$$

$$4:(e)\,xxx;\quad \bar{x}\bar{x}x;\quad \bar{x}x\bar{x};\quad x\bar{x}\bar{x}$$

We place two zinc atoms in positions 1a and 1b, so that they form a body-centered lattice. This distribution agrees with the above conclusion about the distribution of heavy atoms in this structure. The light atoms 4C and 4N, which are indistinguishable in X-ray analysis, were placed in position 4e which will be occupied twice. Hence the structure under investigation has two parameters. If we assign them the values $x_1=0.2$ and $x_2=0.3$, the intensities of the reflexions calculated for this structure will be in good agreement with the experimental values (Table 2). In this table we assume the value of the intensity of 110-line to be 100, as in the experimental data of Hanawalt and others.

Table 2

Experimental and Calculated Intensities of the Lines on the Powder Photograph of $Zn(CN)_2$

| Nos. | *hkl* | $I_{calc}$ | $I_{exp}$ | Nos. | *hkl* | $I_{calc}$ | $I_{exp}$ | Nos. | *hkl* | $I_{calc}$ | $I_{exp}$ |
|---|---|---|---|---|---|---|---|---|---|---|---|
| 1 | 2 | 3 | 4 | 1 | 2 | 3 | 4 | 1 | 2 | 3 | 4 |
| 1 | 100 | 0 | 0 | 11 | 311 | 1.0 | 0 | 21 | 420 | 2.8 | 3 |
| 2 | 110 | 100 | 100 | 12 | 222 | 2.0 | 1 | 22 | 421 | 0.5 | 0 |
| 3 | 111 | 10 | 13 | 13 | 320 | 0 | 0 | 23 | 332 | 1.5 | 1 |
| 4 | 200 | 4.2 | 4 | 14 | 321 | 19 | 20 | 24 | 422 | 3.1 | 1 |
| 5 | 210 | 0 | 0 | 15 | 400 | 2.1 | 1 | 25 | 500 | 0 | 0 |
| 6 | 211 | 33 | 40 | 16 | 410 | 0 | 0 | 26 | 430 | 0 | |
| 7 | 220 | 21 | 27 | 17 | 322 | 0 | | 27 | 510 | 2.8 | 3 |
| 8 | 300 | 0 | 0 | 18 | 411 | 5 | 7 | | | | |
| 9 | 221 | 0.6 | | 19 | 330 | 4.4 | | | | | |
| 10 | 310 | 11 | 11 | 20 | 331 | 0.1 | 0 | | | | |

The intensities were calculated using the formula

$$I_{calc} \approx p\frac{1+\cos^2 2\theta}{\sin^2\theta\cos\theta}|F|^2 e^{-b\left(\frac{\sin\theta}{\lambda}\right)^2}$$

in which $p$ is the number of co-operating planes; $\theta$, the reflexion angle; $F$, the structure amplitude; $\lambda$, the wave length. The exponential function is the temperature factor, in the calculation of which we assumed $b=4$. The atomic scattering factors were taken from the International Tables [2], $f_N=f_C$ being accepted. For the atomic factor of the bivalent Zn-ion that of a nickel atom was taken.

The good agreement between the calculated and the experimental data in Table 2 confirms the validity of the structure chosen. Thus, the structure of $Zn(CN)_2$ is described by means of the following co-ordinates:

$$Zn\left(000,\ \tfrac{1}{2}\tfrac{1}{2}\tfrac{1}{2}\right);\quad 4(C+N)\ \text{ in }\ 4e_1(x_1x_1x_1)+4e(x_2x_2x_2)$$

$$x_1=0.2;\quad x_2=0.3$$

The space group is $T_d^1$ (accepting the highest point symmetry of the atoms). The structure of the $Zn(CN)_2$ is shown in Fig. 1. This structure closely resembles that of anticuprite. The centres of the groups C—N form the face-centered lattice, which is placed in a body-centered lattice of Zn-atoms (the displacement of both lattices along the body diagonal being equal to $\frac{1}{4}$). The distance Zn—C—N—Zn $a\sqrt{\frac{3}{2}}=5.1$ Å is in good agreement with that found in the structures of cyanides with univalent cations [type of the formulae $FeR_n$ $Fe(CN)_6$, where $R$ is an univalent cation, $n=0,1,2$], investigated by Keggin and Miles (3), and in the structures of cyanides with bivalent cations [type of the formulae $M_3[Fe(CN)_6]_2$ where $M$ is a bivalent cation], investigated by Van-Bever (4). In the two last types of cyanides the heavy atoms form the primitive cubic lattice with a unit cube edge 5.1 Å.

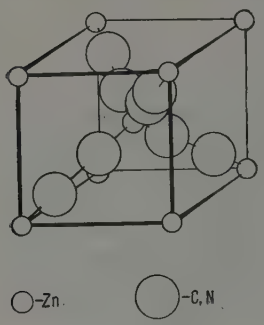


Fig. 1.

It is possible to describe the structure of $Zn(CN)_2$ as a compact cubic packing of cyanide groups, a quarter of the tetrahedral holes of which are filled with zinc atoms.

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*GEOLOGY*

# THICKNESS OF THE RECENT DEPOSITS OF THE VOLGA RIVER DELTA

By M. V. KLENOVA, N. G. KRASNOVA and K. A. RACHKOVSKAYA

*(Communicated by V. A. Obruchev, Member of the Academy, 17. II. 1941)*

Light drilling (to 20 m) with the purpose to elucidate the composition of the subsurface rocks while preparing a lithologo-morphological map, was made by an expedition under the Fish Industry Administration (Glavrybvod), sent out for an aerial survey of the Volga river delta.* A total of 50 holes were drilled, uniformly distributed throughout the whole of the area mapped, i. e. in the delta proper and in the western and eastern steppe lagoons.

The geology of the region investigated has until now been known principally from the works of Pravoslavlev (1) who made observations chiefly on the exposures of the Ber knolls. The drill-holes described by Bogdanov (2) are situated upstream of Astrakhan beyond the boundaries of the region covered by our work. The deep-hole sunk on the Parabichev Knoll on the margin of the town of Astrakhan, has reached 700 m, but in the description of Bogdanov (2) the upper packs of strata to the Baku strata are not sufficiently subdivided. A light drilling (to 10—15 m) had been carried out by the expedition of Zentromorproject in a number of points. Different organizations had also sunk separate shallow drill-holes at some points for building and similar purposes, but all these data have been either lost or cannot be used, owing to a difference in the method of describing and treating the material. Therefore the comparisons below have been made only on the basis of our own data.

The points of the holes were plotted by us on a scheme of the hypsometry of the Volga delta (Fig. 1), compiled by the Laboratory of Hydrology of the Volga-Caspian station, the All-Union Institute of Fisheries and Oceanography, similar to a bathymetrical map, with the use of numerous detached levellings, carried out in the delta by different organizations previous to the works of the Expedition for Aerial Survey.

The scheme shows the conventional isolines on the basis of only the altitude marks of dry land, without taking into consideration the level of water in the river. Such a method of representing the slightly developed relief of the Volga delta brings up the nature of its surface in a descriptive manner.

* The expedition was organized in 1936 in fulfilment of a special order of A. I. Mikoyan, People's Commissar of Fish Industry, and is at present completing the preparation of a topographic map of the delta on a scale 1 : 100000, and of special maps—the morphological, soil and vegetation maps.

In the scheme there are sharply distinguished the areas of the western and eastern steppe lagoons with numerous Ber knolls; clearly seen are also the more elevated massifs on the western part of the delta proper between the Volga and Bolda rivers, and in the eastern part between the system of the Buzan and Churka rivers on the one hand, and of the Krivoy Buzan and Sumintza Shirokaya, on the other. These areas are located within the boundaries of the isoline—24 m from absolute zero. In these areas the Ber knolls are rather

Fig. 1.

sparse, but still widespread enough. Between these more elevated areas in the system of the Dolgaya, Rycha, Bushma, Belesky Bank rivers, there is situated a region where the isoline —25 m enters the middle portion of the delta, and the isoline —24 m rises upward nearly to the upper reach of the Buzan. The Ber knolls here are either lacking altogether, or arranged in separate isolated monadrocks.

The holes studied by us show that the peculiarities of the relief of the Volga delta noted are quite regular and are conformed by its geological structure.

At the majority of points under the recent, mostly alluvial, deposits the holes have exposed older rocks. According to Pravoslavlev, the upper

member of the Quaternary deposits in the delta are the rocks of the Ber knolls, underlain by a characteristic horizon of chocolate Khvalynsk clays. A typical section of such a kind is given by our hole No. 29 near the knoll of the Serghievka village, where from the top down to a depth of 7.82 m lie bright-coloured brown arenaceous clays (light loams) with fragments of sea shells, which downwards are replaced by a thickness of brown (chocolate) clay with a well-developed angular cloddy structure, with interbeds of a more arenaceous material, fragments of *Cardiidae* and *Dreissensia*. At the depth of 10.68 m from the surface the clay is replaced by yellow sands about 9 m thick. Under the sands lies a thickness of mainly clayey rocks, grey in colour, with some intercalations of a more sandy material and with a slight brown tinge in the upper portion.

When arranging the holes according to the hypsometrical scheme, it may be seen that in the higher areas this section is repeated with the exception of the uppermost portion of the knoll rocks, since some holes are sunk in inter-knoll depression. Most persistent throughout the delta is a horizon of yellow sands, lying in hole No. 29 under the Khvalynsk clays.

**Thickness of the Recent Deposits, According to the Regions of the Volga Delta**

| Region | Variation of thickness, m | Average thickness, m | Height, absol. mark, m |
|---|---|---|---|
| Western steppe lagoons | 0 — 6.18 | 2.5 | —21.9 |
| Western elevated area | 0.23— 9.36 | 6.6 | —23.9 |
| Central depression | 6.15—20.36 | 14.0 | —24.3 |
| Eastern elevated area | 7.30—15.07 | 11.9 | —24.1 |
| Eastern steppe lagoons | 0.22— 8.40 | 5.2 | —24.3 |

As it will be seen from the table, the thickness of the recent deposits in the region of the western lagoons is very small—on the average 2.5 m. It is gradually increasing towards the East, reaching a maximum value in the region of the depression, where our holes have not always reached the underlying rocks (therefore the average thickness of 14.0 m is too low). Further East the thickness of the deposits is decreasing again. The absolute heights of the holes are varying very slightly, from—22.40 to—25.70 m, and these variations in every region take place in the direction of its long axis from the upper course of the delta towards the sea, which is reflected also in the average values that are nearly similar in all the regions.

Thus, in the very slightly dissected recent relief of the Volga delta our work has revealed the ancient relief reflected in the varying thickness of the alluvial cover*.

As shown by the observations of the Laboratory of Sea Geology at Astrakhan (M. P. Gudkov), the growth of the Volga delta at present is made chiefly at the expense of the drop in the level of the Caspian Sea. The annual increase in thickness of the spits at the expense of the accumulation of new

* Assuming the area of the western steppe lagoons (of the area mapped) to be 1800 km², of the western elevated region—2100 km², of the depression—3200 km², of the eastern elevated area—2400 km² and of the mapped part of the eastern steppe lagoons—1000 km², and multiplying these figures by the corresponding thickness, we obtain that the total volume of the alluvial deposits of the Volga delta is equal to 96.92 km³; dividing this value by the yearly volume of matter washed away by the Volga—0.013 km³ (³), the approximate age of the present delta is about 7500 years.

alluvium does not exceed 10.8 cm in the most rapidly growing portion of the shore-line of the delta. The increase in the thickness of the deposits in the delta itself takes place at a much slower rate. The relatively insignificant values of thickness obtained correspond to the observations available on a number of other deltas. Thus, in the delta of the Mississippi at New-Orlean the Quaternary clay lies at a depth of 9.5—16 m ([4]). In the delta of the Ganges to the deltaic deposits proper belongs a thickness of sands and clays measuring 18 m on the average. In the delta of the Nile the thickness of the present recent deltaic deposits are underlain by an ancient sea sand, that had covered the bottom of the bay, later on filled by the alluvium of the Nile.

If on the basis of the thickness of the alluvial cover of the Volga delta, observed by us, its age be calculated, we shall obtain a value of the order of 7.5 thousand years, which suits rather well the Quaternary chronology, since 8—10 thousand years ago at the place of the present delta of the Volga a bay of the Khvalynsk Sea was still in existence.

The insignificant value of the average increase in thickness (in the central part it is equal to 2—3 mm per year, according to these calculations) may be well accounted for by the fact that simultaneously with the accumulation of sediments in the delta their intensive erosion is active, accompanied with the formation of new channels, levelling of the ancient relief, etc.

A detailed study of the materials from drill-holes, which is carried on at present, aims at elucidating the Quaternary history of the complex geological phenomenon which we call the delta of the Volga.

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*MINERALOGY*

# A NEW LUMINESCENCE METHOD FOR ESTABLISHING THE BITUMINOUS NATURE OF ROCKS

By V. N. FLOROVSKAYA

*(Communicated by S. I. Wawilow, Member of the Academy, 12. II. 1941)*

The present preliminary note is devoted to results obtained in the study of the luminescence of fossil organic compounds, in particular to the study of the luminescence of bitumens, this problem being of great industrial significance.

The investigation was conducted in the ultra-violet, the radiation being produced by a quartz mercury lamp fitted with a Wood filter ($\lambda = 3000$—4000 Å) and a spark gap with a monochromator ($\lambda = 2200$—4000) designed by the State Optical Institute (1). The fossil organic compounds investigated are arranged in order according to the classification adopted by Dana (2).

a) Among the investigated salts of organic acids (wavellite, oxammite, mellite) only mellite (honey-stone) showed a bright blue-violet luminescence in the short wave ultra-violet.

b) All the resins and the crystalline products of their transformation, exhibited at this Museum (amber, roumanite, walchowite, retinite, ambrite, fichtelite), showed an identical yellow-green luminescence, especially bright in the long wave ultra-violet.

c) Ozokerites from the deposits of the Moldavian Republic, Caucasus and the Island of Cheleken were found to display a yellow-brown luminescence, brightest in the long wave ultra-violet radiation.

d) The investigated liquid bitumens (oils) from some of the deposits of the USSR displayed a luminescence varying in colour, but in all cases brightest in the short wave ultra-violet. Ukhta oil—a yellow-brown luminescence; Emba oil—a yellow-blue luminescence; Azerbaidjan oil—a bright blue luminescence; Balakhany oil—a light blue luminescence.

e) The asphalts of the Syzran region were found to have a dark brown luminescence in the long wave ultra-violet; in the short wave radiation their luminescence was similar to that of amber.

f) The product of the modification of liquid bitumens—elaterite—had a brownish luminescence in the long wave region of the ultra-violet.

g) Canadian albertite, Caucasian anthraxolith, graphite of Ceylon and schungite did not luminesce.

For all the investigated minerals, except the resins and mellite, it was found possible to make the phenomenon of luminescence more vivid both in brightness and in colour. This was achieved by moistening the minerals with some organic solvents (chloroform, petroleum ether, benzine, benzene).

The moistened area of the mineral showed a vivid luminescence, its colour depending on the nature of the bitumen. The observed colours of luminescence conformed with those of the luminescence of oils, given above.

It is noteworthy that the pyro-bitumen albertite, Pikäranta graphite, graphitized bitumen and graphite of the Turukhan region, none of which does generally luminesce, did so when a drop of organic solvent was put upon the surface of each. These observations apparently imply that the method employed by the author can to some extent solve problems as to the origin and genetical relation of the caustobioliths.

With the decrease in the content of bitumen, the usual yellow-brown luminescence of bituminous minerals observed in the long wave ultra-violet light disappears. Consequently, there is no possibility of discerning such rocks by their luminescence. This difficulty is, however, soon overcome by employing the method based on the described above property of bitumens. After a drop of the non-luminescent organic solvent is put on to the surface of the bituminous mineral, the moistened area begins to luminesce. Chloroform was found to be the most suitable solvent for this purpose. Employing the described method, it is possible to detect, without affecting the integrity of the mineral, a bitumen, the content of which is considerably less than 0.01%.

It may be shown without difficulty that the luminescence of the moistened area of the bituminous rock is accounted for by the bitumen contained in the rock. For example, after removing the bitumen from the rock by means of usual extraction, the rock, although moistened with organic solvent, does not luminesce.

The method employed by the author makes it possible to determine the presence of bitumens in sands, sandstones, clays, shales and limestones and roughly to estimate their content. In the case of limestone the surface of the rock must be somewhat loosened by simple scratching.

The rapidity, simplicity and high sensitivity of the described method of identifying bituminous rocks prompted the author to communicate her observations. The described method may be employed directly in places of drilling. All cores may be systematically subjected to tests as to their possible bituminous nature. Moreover, old cores kept at storehouses may also be revised.

The practical utilization of this method will allow the bitumen laboratories to dispense with many useless analysis of non-bituminous rocks.

At the present time the Mining Museum, together with the Luminescence Laboratory of the State Optical Institute, is developing the described method and designing corresponding portable field apparatus. The introduction of the method into the praxis of oil industry is of exceptional interest to day, when the real value of new oil-bearing regions of atypical structure is being investigated.

Mining Museum.
Leningrad.

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*MINERALOGY*

# A LUMINESCENCE METHOD FOR DETECTING URANIUM MINERALS AND ORES

By V. G. MELKOV and Z. M. SVERDLOV

*(Communicated by S. I. Wawilow, Member of the Academy, 8. II. 1941)*

The absence of a speedy and reliable method for the diagnosis of uranium minerals directly in workings makes prospecting for uranium deposits an arduous task. The new luminescence method of ore analysis described below may be employed both in the laboratory and in the field.

Though in the majority of cases the luminescence of inorganic compounds is due to the presence of impurities, the luminescence of the uranyl salts is one of the rare cases of true luminosity, a comprehensive study of which has been made by a number of investigators ([1–3]). The new method is based on this property of uranium.

Two types of commercial deposits of uranium are known, the hydrothermal and the sedimentary. The principal minerals of industrial importance belonging to the hydrothermal deposits of uranium are nasturan and uranium black, the general composition of which may be represented by $mUO_2.nUO_3.qPbO$. Minerals of the zones of oxidation of these deposits, i. e. minerals of the type uranium-mica, the uranyl sulphates and uranyl silicates, generally are of little interest.

The most important minerals belonging to the deposits of the sedimentary type are uranium vanadates of potassium and calcium.

It is an extremely difficult and at times impossible task to detect thin veins of nasturan and yellow, black, grey or ashy accumulations of uranium black. This difficulty is especially aggravated by underground conditions of mining, in which case the diagnosis of dark uranium minerals in situ is particularly essential for documentation.

In mining practice microchemical or drop-reactions and radiometric measurements are employed to determine uranium carrying minerals. All these tests require laboratory conditions and, naturally, can be performed on pieces of ore only, and the quantitative results obtained lack accuracy. The difficulty of diagnosing dark oxygen-containing uranium minerals among other dark accompanying minerals possibly accounts for the small number of uranium deposits of the hydrothermal type known in the USSR.

It is to overcome these difficulties that the authors of this paper have employed the method of luminescence analysis. Sources of short wave and long wave ultra-violet rays were used to excite the luminescence of nasturan and uranium black. However, in neither case, i. e. neither for nasturan nor for uranium black, was the phenomenon of luminescence observed.

The rapid transformation of uranium black and nasturan under natural conditions of the zone of oxidation into vividly luminous compounds, mainly uranyl sulphates, and also the well-known characteristic luminosity of the uranyl salts in solutions ([3]), suggested to the authors a means of converting the non-luminous uranium compounds of the $mUO_2.nUO_3.qPbO$ type into compounds endowed with the property of luminescence.

To obtain luminescent compounds of uranium upon uranium black, nasturan and their ores, the method of spraying these minerals with various solvents was employed. Sulphuric, nitric, hydrochloric, acetic and phosphoric acids of various concentrations were used as solvents. It was found that in every case there were formed compounds of uranium luminescent, preferably under the action of short wave ultra-violet radiation. These compounds showed characteristic band spectra. The compounds obtained were found to be luminous both immediately after the spraying of the mineral under investigation and after its drying. In the case of phosphoric acid being employed as solvent the luminosity was weak. Acetic acid proved to be the most convenient, since this acid dissolved uranium minerals well even in weak concentrations (1—2%). This acid proved poorly active with respect to many mineral components.

The method described is also applicable to uranium vanadates when strong acids are used.

For the excitation of the luminescence of uranium minerals, the authors are perfecting a portable source of ultra-violet radiation of the type of Komovsky ([5]). This apparatus must include a special light filter, which makes the short wave region of the ultra-violet spectrum stand out. A special pulverizer is also being designed by the authors to spray the surface of the mineral with the solvent.

The method of exciting the luminescence for the purpose of diagnosing minerals non-luminous under natural conditions, by converting these minerals into luminous compounds, may well be applied to a number of other natural products.

The obtained results may be used for the practical purposes of diagnosing uranium black and nasturan directly in drifts, workings, as well as in the laboratory.

Mining Institute.
State Optical Institute.
Leningrad.

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*PETROGRAPHY*

# MICACEOUS KIMBERLITES IN THE NORTH OF THE CENTRAL SIBERIA

By GEORGE MOOR

*(Communicated by V. A. Obruchev, Member of the Academy, 5. II. 1941)*

In one of the earlier papers (1) the present author mentioned the fact that he succeeded in establishing in the river Kheta basin the presence of peculiar alkaline gabbroids and their effusive analogues—melilite basalts, resembling the kimberlites of S. Africa. In this connexion he suggested the possibility of diamond occurrence in the northern parts of Central Siberia. At the present time some new data give grounds for speaking with greater certitude of the presence of magmatic formations within the limits of northern outskirts of the Middle Siberian Plateau and Taimyr and of similarity of these formations to micaceous kimberlites of the South African Platform area. This new factual evidence is furnished by a rock of the alneite type, a sample of which was brought by the Taimyr geological expedition organized in 1929 by the Geological Committee with the purpose of crossing the Central Taimyr.

V. Sobolev (2) was the first to study this sample in thin-sections and to point out that its composition is determined by a combination of a densely coloured zonal biotite, melanite (titanoferous garnet) also zonal, augite, apatite, pseudomorphoses on olivine, and of a vitreous basis which apparently represents an aggregate of ceolites. Biotite builds up the porphyritic excretions and forms more or less idiomorphic tablets. It is characterized by a zonal distribution of colouring and closely resembles the biotite of peridotites and of some lamprophyres. The external portions of biotite plates exhibit a reverse (tourmaline) scheme of absorption, the colour varying here from greenish to orange; but in the central parts of plates the normal scheme of absorption is retained, and the colour changes from yellowish to light orange. V. Sobolev indicated that the mineralogical composition of this rock places it closely to the group of alneites or effusive lamprophyres. On the same occasion this author pointed out that although the rock described above is probably of hybrid nature, the possibility of existence of a new heretofore unknown alkaline province in the North of the Central Siberia is not excluded.

While studying the rock from the river Upper Taimyra in additional thin-sections obtained from more fresh portions of the sample, the author of the present paper ascertained, besides the above-mentioned minerals, the presence of small quantities of nepheline, which is strongly decomposed and nearly completely converted into an aggregate of ceolites, calcite, scapolite.

This aggregate contains, as may be seen from the value of birefringence, a considerable amount of the meyonite molecule, epidote and perovskite. For biotite $Np' = 1.596 \pm 0.004$; $Nm = 1.612 \pm 0.002$ (sometimes it drops to $1.608 \pm \pm 0.002$) and $Ng' = 1.636 \pm 0.004$.

The complex of minerals enumerated above, together with large porphyritic excretions of biotite, are responsible for the similarity of the rock from the river Upper Taimyra to the micaceous kimberlites of South Africa. The latter, according to P. Wagner ([3]), consist of crystals of olivine and phlogopite immersed in a basic mass, composed of small grains of mica, augite, apatite, perovskite and ferruginous ores; besides there are also present small quantities of serpentine and calcite, which are probably the products of alteration of glass. These rocks are distinguished from basalt-kimberlites by a lower magnesium content and a higher content of sodium, but display in this respect a marked similitude with lamprophyres and melilite basalts.

The chemical composition of micaceous kimberlite from the river Upper Taimyra is given in the following table.

| $SiO_2$ | TiO | $Al_2O_3$ | $Fe_2O_3$ | FeO | MnO | MgO | CaO | $K_2O$ | $Na_2O$ | $H_2O > 110°$ | $H_2O < 110°$ | BaO | Total |
|---|---|---|---|---|---|---|---|---|---|---|---|---|---|
| 34.81 | 2.40 | 12.59 | 7.17 | 5.28 | 0.15 | 15.58 | 11.17 | 3.14 | 0.68 | 4.88 | 1.48 | 0.42 | 99.65 |

Alneite or micaceous kimberlite was found on the river Upper Taimyra, above the issue of its large right tributary, the river Gorbita. The river Upper Taimyra flows here among Quarternary deposits of the Taimyr depression, from under the eroded surface of which come out at places the underlying Upper Paleozoic deposits. The micaceous kimberlite was discovered in a bedrock among a comparatively small isolated outcrop of carbonaceous and carbono-argillaceous shales which are likely to belong to the productivity deposits referred on the Taimyr by most investigators to the Upper Permian. Shales at the contact are partly fused and injected, this creating a picture of a peculiar intrusive breccia. The presence of altered argillaceous shales is representative of the fact that the micaceous kimberlite is younger than the Upper Permian. A number of inderect evidences, based in the most part on data of the geology of islands of the Central Polar Basin (Bennet island, etc.) may lead one to the belief that the micaceous kimberlite belongs to the Upper Mesozoic or even to the Tertiary. It should be borne in mind, however, that such assumptions require further proofs.

In this connexion it may be mentioned that this assumption has been challenged by Saks in one of his last papers on the tectonics of the Taimyr depression ([4]). He remarks that if the intrusion of alkaline rocks were dated to the Upper Mesozoic or to the beginning of Tertiary, their formation would be associated with that of the Taimyr depression, and outcrops of alkaline rocks would have been found among Mesozoic and, in particular, Jurassic and Cretaceous deposits, which are widely developed in the eastern part of the depression. Now as no magmatic formations are known in the region of the development of Mesozoic within the boundaries of the Taimyr depression, he believes it more correct to assume that the formation of alkaline rocks preceded the deposition of the marine Mesozoic in the Taimyr depression, or, in other words, occurred prior to the Jurassic. This assumption is not ungrounded and should be taken into account when evaluating the age of alkaline rocks.

On the other hand, our present knowledge of the Taimyr depression is not sufficient to give the degree of finality to the assertion that there exist no

evidences of magmatic activity in the region of distribution of Mesozoic beds of the Taimyr depression. It is universally known that alkaline rocks form, as a rule, small outcrops which may be detected only in the case of detailed investigations, which up to the present time covered but small areas within the boundaries of the Taimyr depression.

A chemical analysis of the rock from the river Upper Taimyra shows its chemical peculiarities to be very similar to those of micaceous kimberlites and melilite besalts from northern outskirts of the Siberian platform (basin of the river Medvejaya). Respective values are set out in the table*.

| Nos. | $SiO_2$ | $TiO_2$ | $Al_2O_3$ | $Fe_2O_3$ | FeO | MnO | MgO | CaO | $K_2O$ | $Na_2O$ | $H_2O>110°$ | $H_2O<110°$ | Total |
|---|---|---|---|---|---|---|---|---|---|---|---|---|---|
| 1 | 34.81 | 2.40 | 12.59 | 7.17 | 5.28 | 0.15 | 15.58 | 11.17 | 3.14 | 0.68 | 4.88 | 3.48 | 99.65 |
| 2 | 31.80 | 1.40 | 3.41 | 5.19 | 3.47 | — | 24.69 | 10.04 | 4.38 | 0.29 | 5.59 | 0.63 | 99.98 |
| 3 | 31.17 | 2.96 | 5.35 | 4.78 | 9.76 | — | 19.90 | 17.76 | 2.51 | 2.03 | 2.05 | 0.44 | 99.62 |
| 4 | 39.02 | 3 70 | 3.81 | 7.69 | 9.72 | 0.20 | 19.47 | 11.20 | 0.92 | 0.40 | 3.53 | 0.76 | 100.42 |

The first of these analyses refers to the micaceous kimberlite from the river Upper Taimyra; the second and third—to micaceous kimberlites and melilite basalts of South Africa; and, finally. the fourth—to melilite basalts of the river Medvejaya basin.

A comparison of these analyses shows that the rock from the river Upper Taimyra most closely resembles the micaceous kimberlites of South Africa. As far as melilite basalts from the river Medvejaya basin are concerned, they may be compared, judging by their chemical composition and presence of magnesial olivine in inclusions, with basalt-kimberlites. Therefore within the boundaries of the northern parts of the Central Siberia there can be notified rocks of both micaceous kimberlite and basalt-kimberlite types. This fact enables a comparison of the northern parts of the Central Siberia with the South African Platform to be done with greater certitude and shows the possibility of the former being diamond bearing. In this connexion it seems desirable to carry out a number of geological investigations with the purpose of detecting bedrock outcrops of kimberlite rocks within the limits of northern parts of the Siberian Platform and Taimyr, and also to start washing of friable materials in order to ascertain the supposed presence of diamonds.

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*Translated by P. Rappoport.*

* The first and forth analyses have been carried out in the Geochemical Laboratory of the Chief Administration of the Northern Sea Route (Glavsevmorput) (analysts E. Tolkacheva and A. Lubelskaya). The second and third are derived from P. Wagner's paper [3]. In the first analysis is not included BaO content —0.42%, and in the second and third $P_2O_5$ content, equalling 1.49 and 1.69%, respectively. Besides, the $CO_2$ content, 7.65%, is omitted from the analysis of micaceous kimberlite.

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*STRATIGRAPHY*

# OCCURRENCE OF UPPER PERMIAN PELECYPODS IN WESTERN TAIMYR

By E. M. LUTKEVICH

*(Communicated by A. A. Borissiak, Member of the Academy, 1. III. 1941)*

During the last years there has been collected from the deposits of the Tungus series of Western Taimyr a fauna, according to which these beds may be subdivided into deposits of the Upper Devonian, Lower Carboniferous and Lower Permian ([1],[2]). In all of the investigated localities of the Taimyr peninsula the Lower Permian deposits are underlying coal-bearing deposits, where collections of flora and pelecypod fauna have been made. The flora was partly described and determined by E. Chirkova ([18]), N. Shvedov ([14]) and G. Radchenko ([7]). These authors, upon comparing it with the flora of the Kuznetsk basin, regard the lower part of coal-bearing deposits of Taimyr as Lower Permian, and the upper one, of considerably less thickness, as Upper Permian. The pelecypod fauna collected by the geologists who worked on the Taimyr, and by the author, is met almost in all the Permian deposits, from the Lower to the Upper ones inclusively, represented in purely marine, littoral coal-bearing and continental facies. At the present the author is preparing a monograph on the fauna, but the major part of specific determinations has already been made, thus enabling us to establish the age of the coal-bearing deposits of the Taimyr as Upper Permian.

The brachiopod fauna from the Lower Permian deposits of Taimyr has been determined by B. Licharew, A. Rotai and O. Einor ([1]). Papers published during the last years, describing brachiopods from the Permian deposits of the Arctic region of the USSR ([4], [15], [19]), report the occurrence of Upper Permian forms of brachiopods in association with fauna from the Lower Permian, still bearing relation to the fauna of the Carboniferous period, if one is to judge by a number of species of wide vertical distribution. The following pelecypods have been listed from the Lower Permian deposits of the Western Taimyr, characterized by brachiopods: *Nucula wymensis* Keys., *Nuculana speluncaria* Geinitz., *Astartella permocarbonica* (Tschern.), *Allorisma elegans* King., *Edmondia murchisoni* King., *Pecten (Pseudamusium) pusillus* (Schloth.), *Aviculopecten serdobowae* (Fred) and a number of forms, belonging to the genera *Modiola* and *Allorisma*, not yet specified. None of the species enumerated is met below the deposits of the Artinskian stage, but they are all widespread in the Kazan stage and zechstein of Western Europe, except the last-mentioned species.

Thus, the pelecypod fauna, as well as the brachiopod one from deposits referred to the Lower Permian on the Taimyr and underlying the coal-bearing deposits, contain a considerable number of Upper Permian species. The question arose a few years ago as to what had been the course of migration of the Upper Permian marine fauna to the Russian platform, and after the

appearance of papers treating of brachiopods of the Arctic, it became clear that migration proceeded from the North. However, the question as to the exact time of the migration of the fauna is still undecided for lack of correlation of the deposits of the Kazan stage with the marine deposits of the Permian of the Arctic. With the present subdivision of deposits of the Lower Permian of the Arctic it is too early to assign part of them to the Upper Permian, no sufficient reasons having been advanced to support this contention. Hence deposits with brachiopod and above-mentioned pelecypod fauna have been on the Taimyr, too, referred to the undivided Lower Permian of about 1200 m in thickness ([1],[2]).

These deposits are overlain by strata with rare coal seams, a still sparse flora and a distinct pelecypod fauna making their appearance. Of brachiopods these deposits have yielded but *Lingula credneri* Geinitz. var. nov. and *Rhynchopora* cf. *geinitziana* Vern., determined by O. Einor. From the same lower beds of the coal-bearing deposits have been identified: *Pecten* (*Pseudamusium*) *sericens* (Vern.) of the genus *Nuculana*, *Allorisma* and *Edmondia*,

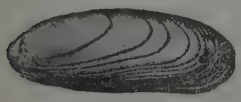

Fig. 1. *Netschajewia pallasi* (Vern.) with *Spirorbis* seated on the posterior edge. Left valve. Matveevka rivulet, 5 km from the coast of the Yenissei bay. Natural size. Near the anterior end is seen the impression of the anterior adductor, separated by a furrow.

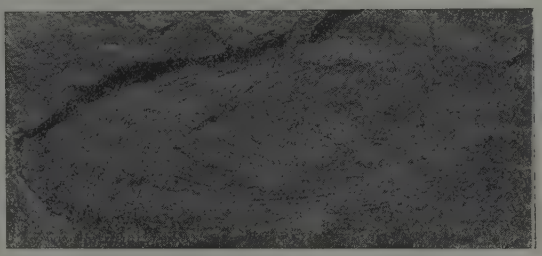

Fig. 2. *Procopiewskia gigantea* Ragos. Right valve. Matveevka rivulet, 8 km from the coast of the Yenissei bay. $^2/_3$ of natural size.

not yet classed as species, some forms not having been yet generically determined. It seems that our present knowledge of the pelecypod fauna and the identification of but two species of Upper Permian brachiopods still furnish evidence for referring these deposits to the Upper Permian and separating them from the Lower Permian of Taimyr. Besides, the peculiar composition of this fauna, coupled with its occurrence in the base of the coal-bearing deposits, and the lack of forms in common with the subjacent deposits, give us ground for assigning it to the Upper Permian. This is exactly what has been done by establishing the Piassina series, up to 1000 m thick ([1],[2]).

In the overlying deposits coal beds occur more frequently, accumulations of flora become more abundant and an exclusively pelecypod fauna appears in large numbers of specimens. From their list are worth noting species already described, whose range of distribution is known. The following are the species concerned: *Netschajewia pallasi* (Vern.) (Fig. 1), *Netsch. tschernyschewi* Lichar., *Netsch. globosa* Netsch., *Netsch. oblonga* Golowk. and a number of new species of this genus; *Pleurophorus costatus* Brown., *Pl. oblongus* Meek and Hayd., *Pl. subcuneatus* Meek and Hayd., *Liebea hausmani* (Goldf.), *Liebea septifer* King., *Nuculana kasanensis* Vern. and new species of the genera *Edmondia*, *Parallelodon* and *Myalina*. Part of forms from these deposits proved to be close to those described from the Kuznetsk basin under the generic name *Anthraconauta* ([10],[11]). Among them one species ([5]) can be identified with precision as *Procopiewskia gigantea* Rag s. (Fig. 2), whose forms have been found in association with the enumerated species both in one layer and

throughout the entire series of coal-bearing deposits, separated under the name of the Lemberov series, attaining a thickness of more than 1000 m ([1,2]). Besides these forms, great development is attained by forms of the new genus *Taimyria*, comprising several species and characterized by the presence of a keel on the shell carrying dichotomous concentric ribs. As to the occurrence of *Procopiewskia gigantea* Ragos., widespread in the deposits of the tops of the Balakhonka and Bezugolnaia series of the Kuznetsk basin, it points to the possibility of accurate comparison of the deposits of Taimyr and Kuznetsk basin. However, it furnishes no basis for determination of the geological age, since that of the coal-bearing deposits of the Kuznetsk basin is not to be considered as established, judging by its various determinations, according to fauna and flora ranging from the Lower Carboniferous to the Upper Permian.

Owing to the presence on the Taimyr of forms characteristic of the Kazan stage of the Russian platform and those from the uppermost parts of the Balakhonka and Bezugolnaia series of the Kuznetsk basin, it may be possible to regard the fauna of coal-bearing deposits of the latter not as an endemic one, but rather as one which during certain periods of the Permian stood in some connexion with the neighbouring regions. Of fresh-water forms has been found but *Najadites* ex gr. *flexuosa* Dix and Truem nov. sp. ([17]) in a stratum crowded with plant remains in the top of the lower part of the coal-bearing deposits. The Taimyr pelecypods may be correlated also with the Gondwana forms as compared to the African ones, the latter including such common species as *Modiolopsis stocleyi* Cox ([16]).

The lowermost parts of the coal-bearing deposits, where fauna of the Kazan stage is developed, attain a thickness of 2000—2500 m. The overlying deposits, where thick conglomerates and polymict sandstones have appeared, are likewise coal-bearing, with a high coal content and lack of fossils. Flora from these deposits, occurring abundantly, especially in the upper part, points to the Upper Permian ([7,13,14,18]). The entire upper part of the coal-bearing deposits has been separated as the Slobodskaya series ([1,2]), according to the characteristic occurrence in the conglomerates developed throughout its thickness of pebblestone, bearing evidence of intensive erosion north of the wash-out region of the Taimyr geosyncline ([3]). Directly beneath the lower cover of effusives, resting upon the upper beds of the Slobodskaya series, on the coast of the Yenissei bay, a pelecypod fauna has been found, as well as in the tuffites, overlying the first cover of effusives. The fauna from these localities proved to be of uniform composition and at the same time identical with the fauna of the Kolchugino series of the Kuznetsk basin. Here have been determined: *Posidonomya subovata* Jones, *Palaeanodonta castor* Eischw., *Palaeomutela* (?) *concina* Jones, *Pal. nanella* Chalf., *Anthracomya fedotovi* Chalf., *Anthr. lata* Chalf., *Anthr. tomiensis* Chalf., *Anthr. simplex* Chalf., *Anthr. subparallela* Chalf., *Anthr. similis* Chalf., *Anthracomya (Anthraconauta) tschernyschewi* Chalf., *Anthr. (Anthraconauta) suprahillipsi* Chalf., *Anthr. (Anthraconauta) porrecta* Chalf., *Anthr. (Anthraconauta) obliqua* Chalf. and a small number of new species of *Anthracomya*. This list of forms determined according to the latest work on pelecypods of the Kolchugino series of the Kuznetsk basin by L. Chalfin, not only shows very close affinity between these forms, but permits to advance stronger arguments when attempting to determine the age of the deposits carrying them. This fauna occurs in the section of the Taimyr deposits, 1500—2000 m above the beds containing the fauna of the Kazan stage, characterizing the time that had passed after the termination of processes of coal accumulation, followed the Pfalz phase of the Hercynian tectogenesis. Consequently, in time this fauna corresponds to the second half of the Upper Permian, which permits to consider this period on the Taimyr as synchronous with the Tar-

tar age. If one should correlate deposits containing this fauna of Taimyr with fauna from the Kolchugino series of Kuznetsk basin, then besides pelecypods, partly pointing to the Tartar age ([12]), from the Kolchugino series have been described by T. Spizharsky ostracods, likewise correlative with ostracods from Tartar stage of the Russian platform ([8]). Hence, according to the position in section of the Taimyr deposits, and on comparing with the ostracod and pelecypod fauna of the Kolchugino series of Kuznetsk basin, there is sufficient ground for synchronizing the deposits containing this fauna with the Tartar stage of the Upper Permian of the Russian platform.

The connexion between the Upper Permian deposits of Siberia and the European part of the USSR, being established on the basis of the pelecypod fauna, furnishes interesting material for drawing a paleogeographic treatment of the Permian time on the territory of the USSR. While on the Russian platform there existed conditions of the epicontinental basin of the Kazan age, succeeded by continental conditions of the Tartar age, the Siberian platform with adjoining regions of geosynclines and regional continental depressions passed through a stage when the latter were filling with thick sedimentation, which points to the vast amplitude of the subsidences of depressions and uplifts of the regions, whence the material was washed down. The paths of migrations and connexions of faunas of these basins led through the Arctic geosynclinal basin, so that synchronous and identical faunas of the Upper Permian deposits linking them together are likely to be discovered on the Pechora and Paikhoe.

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*BIOPHYSIQUE*

# SUR LA DÉFINITION DES VIBRATIONS AÉRIENNES DE FRÉQUENCE ULTRA-SONORE PAR LA MESURE DES SEUILS OPTIQUES

Par K. C. KEKTCHÉEFF et E. P. OSTROVSKY

*(Présenté par P. P. Lasareff, de l'Académie, le 10. II. 1941)*

Au cours des dernières années il a été démontré ([1–4]), que l'action des divers excitateurs sur les sens est accompagnée non seulement de la naissance d'une perception correspondante, mais encore de la variation des seuils d'autres sens, de la pression intraoculaire, de changement du pouls, de la chronaxie motrice, parfois même par l'échange d'énergie, etc. Ce phénomène est appelé action inadéquate des excitateurs. Il est produit, probablement, par le système nerveux végétatif et n'est qu'un reflexe végétatif particulier, causé par l'excitation d'un sens quelconque.

L'un des auteurs a montré précédemment que cette action inadéquate des excitateurs peut ne pas être liée à la perception adéquate. Dans nos expériences nous avons étudié la hauteur des seuils de la vision achromatique, définie par le système nerveux végétatif, en prenant comme excitateur l'énergie rayonnante de longueurs d'ondes différentes, situées des deux côtés du spectre visible. Comme nous l'avons supposé, l'action de ces excitateurs n'est pas accompagnée de perception, mais produit tout de même le déplacement des seuils optiques ([5,6]). On a trouvé la même chose pour les rayons X ($\lambda = 1 \times 10^{-7} - 1 \times 10^{-9}$ cm), pour la radiation ultra-violette ($\lambda = 1 \times 10^{-5} - -1 \times 10^{-6}$ cm) (d'un côté du spectre visible), pour les ondes ultra-courtes ($\lambda = 1 \times 10^{2}$ cm) et courtes ($\lambda = 10^{3}$ cm) (de l'autre côté du spectre).

Ceci laisse à supposer que la partie du système nerveux de l'homme, plus ancienne au point de vue phylogénétique, peut réagir au diapason d'excitateur plus grand (aux ondes électromagnétiques, dans notre cas), que ces excitateurs, dont l'action dépend de la naissance des perceptions subjectives; nous rapportons ces dernières à l'activité de l'écorce cérébrale. Si la conclusion faite est exacte, on pourrait, avec certaine probabilité, s'attendre que les vibrations sonores, ayant la fréquence supérieure à celle du seuil (20000 Hz), puissent également causer une action inadéquate sans produire une perception subjective.

Dans la première série d'expériences nous avons pris, comme source sonore, un sifflet d'Halton. Comme cet appareil produit parfois, au surplus des vibrations sonores non audibles, un faible son perceptible, ce son (mais sans vibrations ultra-sonores) était reproduit par un autre appareil, très primitif. Les déplacements des seuils de la vision achromatique faisant preuve de l'action inadéquate des excitateurs, à savoir, vibrations sonores à la fré-

quence 22000 Hz, étaient obtenus infailliblement dans toutes les expériences avec le sifflet d'Halton*. Le tableau 1 en donne des exemples.

Tableau 1

| L'ob et examiné et la hauteur du seuil / La sensibilité | P-va, 538 22100 | V-ko, 568 22100 | P-va 564 22100 | Tch-va 537 22100 | Tch-va 537 24800 |
|---|---|---|---|---|---|
| Avant le résonnement du sifflet . . | 7.4 | 5.4 | 5.4 | 4.9 | 5.3 |
| Après » » » » . . | 8.4 | 6.6 | 6.5 | 4.2 | 4.4 |

Dans la seconde série d'expériences nous nous sommes servis, pour la production des vibrations ultra-sonores, du principe de magnétostriction. L'appareil correspondant a été construit par le physicien E. P. Ostrovsky (section de physiologie de l'ouïe, directeur—prof. L. A. Andreeff), qui a pris une part active dans toutes les expériences.

Les émetteurs à magnétostriction en nickel étaient excités au moyen d'une installation génératrice munie d'une lampe Г-К-3000, et produisant les vibrations de fréquence 32800 Hz. Les vibrations aériennes de cette fréquence ne sont plus entendues par l'oreille. Pour cela, afin de déceler la résonnance, pour laquelle l'amplitude des vibrations soit maximum, on plaçait sur le diaphragme du tube «sonore» en nickel une goutte d'alcool. Au moment de résonnance il se levait au-dessus du diaphragme un nuage de gouttes microscopiques d'alcool. Le même résultat est observé avec la poudre de lycopode ou du sable, qui commence à s'agiter en sautillant.

Pour la mesure des fréquences acoustiques des vibrations on se servait de la méthode des figures en poussière de Kundt: on obtenait dans un tube spécial en verre des ondes stationnaires produites par l'émetteur en nickel. La distance entre les noeuds était mesurée à 1—1.5% près. En même temps l'on mesurait la fréquence du générateur au moyen d'un ondemètre. Les résultats des deux mesures concordaient à 2—3% près. A la fréquence de 32800 Hz, l'amplitude se trouvait égale à 10 microns.

Les expériences ont été effectuées sur trois objets. En tout, on a fait 57 épreuves, dont 18 de contrôle. Ces dernières reproduisaient entièrement les conditions des expériences fondamentales, mais le tube en nickel était remplacé par celui de laiton. Par suite, la génération des vibrations ultrasonores n'avait pas lieu.

Après que la sensibilité des centres optiques se stabilise à un certain niveau, la personne examinée s'assied à 60—70 cm du générateur. La durée de la génération des vibrations variait de 2 à 3 minutes. Après ceci, pendant plusieurs minutes l'on mesurait les seuils optiques. La majeure partie d'épreuves a donné une élévation de la sensibilité de la vision achromatique (les seuils baissaient).

Tableau 2

| L'objet examiné et le No. de l'expérience / La sensibilité | B-ssof 806a | Tch-va 823a | Tch-va 823a | Tch-va 829 | P-va 817a |
|---|---|---|---|---|---|
| Avant la génération des vibrations . . . . | 2.1 | 4.9 | 4.9 | 5.3 | 5.6 |
| Après » » » » . . . . | 3.0 | 6.2 | 6.3 | 6.2 | 6.4 |

* La mesure des seuils était effectuée par A. V. Kavtorina et A. P. Anissimova.

Si l'on bouchait les oreilles avec de la ouate, les déplacements diminuaient ou disparaissaient tout-à-fait. Ainsi l'on a trouvé, que les vibrations aériennes à la fréquence 32000 Hz, non perceptibles à l'oreil humain, produisent des déplacements de la sensibilité de centres optiques du cerveau.

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*BIOCHEMISTRY*

# ON THE RÔLE OF MALIC ACID IN THE METABOLISM OF THE VEGETABLE CELL

By V. O. TAUSSON

*(Communicated by A. N. Bach, Member of the Academy, 14. II. 1941)*

Organic acids are widely disseminated among the metabolism products of higher and lower plants ([1-6]). This is particularly so in the case of dibasic acids with 4 carbon atoms—succinic, fumaric, tartaric and malic acids. As is well known, the latter is found very often in higher plants, in particular in *Crassulaceae*. In these plants the content of malic acid is subjected to regular daily fluctuations.

Notwithstanding the large number of investigations and of theories that have been forwarded, the physiological rôle and the significance of organic acids, particularly of dibasic acids containing 4 carbon atoms, in the metabolism of the vegetable cell is far from being clear. The majority of old theories, as well as of new ones, dealing with that problem, consider these acids, as well as the other ones, either as the products of disintegration and incomplete oxidation of carbohydrates and aminoacids, or as by-products in the synthesis of proteins. Their further destiny is usually represented as complete oxidation, in the process of respiration, into $CO_2$ and $H_2O$. But there also has been expressed an assumption that they are reduced into carbohydrates([4]).

From the point of view of the theory of the exothermity of biological syntheses advanced by me, it seems hardly possible to agree with such assumption as to the origin, physiological rôle and destiny of organic acids in the living cell. The main thesis of this theory and the issuing consequences ([7]) point at the very important rôle of organic acids, particularly of $C_4$-dicarboxylic acids, in synthetical processes. Corroborative evidence to this statement is furnished by an array of facts: the influence of succinic, fumaric, malic and oxalacetic acids ([8]), as well as of citric acid ([9]), on the oxidative processes in the living cell; the participation of fumaric and oxalacetic acids in the synthesis of aspartic acid in *B. fluorescens liquefaciens* and in higher plants ([10]), as well as in the fixation of elementary nitrogen by nodule bacteria ([11]); and finally, the importance of dicarboxylic acids, oxalacetic and α-ketoglutaric ones, in particular, in the synthesis of aminoacids through transamination ([12-16]).

All of these facts induced me to investigate the nutrient value of $C_4$-dicarboxylic acids, in the first instance of the malic acid, in order to establish whether or not the synthesis of cellular substances is possible by their means. Another aim I had in mind, was to show by the use of a simpler and clearer

method, compared to those employed previously ([7]), that the carboxylic groups of organic acids were not reduced by typical heterotrophic organisms. This would be a further experimental confirmation to the exothermal theory of biosyntheses, and, at the same time, would help the solution of the main problem, which is to find out the significance of malic and of other $C_4$-dicarboxylic acids in the metabolism of the vegetable cell.

The present investigation was carried out by using methods of bioenergetics, by drawing energy balances and determining the coefficients of the consumption of energy* of these acids, as well as of some other substances of similar structure, provided as only sources of carbon ([17–19]).

In the solution of this problem I started from the following conceptions.

The correlation between energy consumption coefficients for malic acid (COOH. $CH_2$.CHOH.COOH) and for substances having a similar structure of the carbon chain and differing only in the number of carboxylic groups [lactic acid ($H.CH_2$.CHOH.COOH) and ethyl alcohol ($H.CH_2$.CHOH.H)] should be as follows:

1) If carboxyl groups are reduced in the processes of heterotrophic synthesis, the maximum coefficient of energy consumption should be provided by ethyl alcohol, and the minimum one—by malic acid. Lactic acid is intermediate. This is easy to understand: as the reduction of the carboxyl groups can take place only with absorption of energy, the additional expenditure of energy (and of matter) will be the larger, the more carboxyl groups are contained in the molecule of the nutrient substance (the structure of the remaining part of the carbon chain being the same). Therefore in such a case the energy consumption coefficient should increase with the decrease of the number of carboxyl groups.

2) If the carboxyl groups are not reduced, but are splitted off as $CO_2$, and if malic acid, as a $C_4$-carboxylic acid, does not occupy a special place in the metabolism of the vegetable cell, then all of the three substances mentioned above should have a nearly equal coefficient of energy consumption.

3) And finally, if the carboxyl groups are not reduced, but malic acid plays a special rôle in synthetic processes and is used as such with its carboxyl groups (i. e. as a dibasic acid), its coefficient of energy consumption should be higher than those of lactic acid and ethyl alcohol. As for the two latter substances, their coefficients are likely to be equal, or nearly so.

These three alternatives made it possible to solve the problem, using experimental data on the coefficients of energy consumption of ethyl alcohol, lactic, malic, succinic, fumaric and tartaric acids.

The necessity of drawing the energy balance determined to a considerable degree the choice of vegetable organisms, the most convenient objects for an investigation of this kind being the fungi. For the experiments described in this paper the following fungi were used: *Aspergillus niger*, *Aspergillus oryzae* and *Penicillium* sp. $Ad_1$, all of them differing considerably by their physiological properties. In order to obtain comparative results, the experiments were carried out using the method of «retarded cultures» ([20]) at 29—30°, as described previously ([7]). The daily dose of the substances mentioned above, which were the only sources of carbon, was 0.001 mole**. The mineral medium was the same as that used in earlier experiments ([7]). Only in the case of *Asp. niger* per every 100 $cm^3$ 0.025 g KCl and 0.0005 g $ZnSO_4$ were added.

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* Ratio between the amounts of energy stored in the organism and that used up in the process of its development.

** In the cultures with ethyl alcohol the syphon pipettes were somewhat modified and provided with a device that warranted the impossibility of losses owing to evaporation.

The methods used for determining the yield and the energy, as well as those used for determining the coefficients of energy consumption, were as reported in ([7],[20]). The main results obtained (average from two parallel tests) are given in the table.

Energy Consumption Coefficients for Ethyl Alcohol, Lactic, Malic, Succinic, Fumaric, Tartaric Acids and for Glucose Determined in the Cultures *Aspergillus niger*, *Aspergillus oryzae* and *Penicillium* sp. $Ad_1$

| Source of carbon | Specific heat of carbon sources, kg cal/mol | Coefficients of the consumption of energy (%) in the cultures | | | | | |
|---|---|---|---|---|---|---|---|
| | | *Asp. niger* | | *Asp. oryzae* | | *Penicillium* sp. $Ad_1$ | |
| | | 9 days | 17 days | 10 days | 18 days | 10 days | 18 days |
| Ethyl alcohol | 329.4 | — | — | 34.1 | 31.4 | 40.1 | 35.1 |
| Lactic acid | 325.9 | — | 29.2 | 34.6 | 33.2 | 40.0 | 35.9 |
| Malic acid | 320.1 | 52.3 | 42.3 | 45.0 | 41.4 | 55.0 | 51.4 |
| Succinic acid | 357.1 | 52.8* | 43.7 | 52.3 | 44.8 | 54.4 | 49.7 |
| Fumaric acid | 320.3 | — | — | — | — | 50.9 | 42.6 |
| Tartaric acid | 275.3 | 48.4 | 42.0 | 41.4 | 38.1 | — | — |
| Glucose | 674.0 | — | — | 42.0 | 37.1 | — | — |

It may be seen from this table that in all cases, notwithstanding the considerable difference in the physiological properties of the three organisms investigated, the energy consumption of malic acid was considerably higher than those of lactic acid and ethyl alcohol. The latter at the same time were very near to each other for every fungus (*Asp. oryzae* and *Penicillium* sp. $Ad_1$).

These very distinct and definite results fully correspond to the last of the three possibilities mentioned above. Therefore, we have the right to draw the conclusion that the carboxyl groups are not reduced in the heterotrophic synthesis, and that malic acid plays a special and prominent part in the metabolism of the vegetable cell, particularly in its synthesis processes. This appears to be the only possible interpretation of the results obtained.

All that has been said above is equally right in respect of other $C_4$-dicarboxylic acids investigated here—succinic, fumaric and tartaric. The reasons for the differences observed in the magnitude of energy consumption coefficients for different dibasic acids in the case of the same organism, as well as for the same acid in the case of different fungi, are due no doubt to the difference in the structure of the carbon chains of these acids, and in a greater degree, to the differences of physiological and biochemical properties of these organisms. Some of them will find their explanation in the results of subsequent studies along the same line.

When considering the data tabulated above, the following circumstances attract attention. The energy consumption coefficient of malic acid (and of other $C_4$-dicarboxylic acids) exceeds considerably (by 10—15%) these coefficients for lactic acid and ethyl alcohol, as well as for glucose, though in the latter case this excess is not so considerable. This makes us suppose that the importance of $C_4$-dicarboxylic acids in the metabolism of the cell is far from being restricted to their catalytic rôles in the intercellular oxidation-reduction processes according to Szent-Györgyi ([8]), and in the processes of amino-acid synthesis through transamination ([12], [13]), as well as to their participation in the synthesis of aspartic acid ([10], [11]). The amount of these acids necessary for carrying out the reactions mentioned above is comparatively small.

* In 13-day cultures.

Therefore, if the rôle of dibasic acids was restricted only to their participation in these processes, their introduction into fungi cultures, as the only source of carbon, could not have influenced the total energy balance in the cell materially and account for such a considerable increase in the energy consumption coefficients of these acids, as compared with those of other substances investigated here.

The existing relations between these coefficients lead us to the conclusion that malic acid, as well as the other $C_4$-dicarboxylic acids investigated, besides acting as catalysts, have a very important influence on the synthetical process, participating actively in the building up of the cell substances with longer carbon chains, probably higher fatty acids and carbon skeletons of aminoacids.

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*CYTOLOGY*

# THE KARYOTYPE OF *RHESUS MACACUS*

By M. I. SOROKINA

*(Communicated by I. I. Schmalhausen, Member of the Academy, 18. II. 1941)*

In spite of the fact that monkeys belong to a rather interesting group of animals, their karyology is studied insufficiently. Only two papers are on hand, that of Painter (1924), and a preliminary note by Shiwago (1939). Painter reported some data on the karyotype of two monkey species: *Brown cebus* (a South American broadnosed species from the *Platyrrhina*), and *Rhesus macacus* (a narrownosed species from the *Cattarrhina*). Spermatogenesis has been studied in both species; moreover, in the case of *Rhesus* Painter investigated the amnion from male and female embryos. Painter considers the number of 54 chromosomes to be typical of the *Brown cebus* monkey, among these is a heteromorphous pair consisting of a larger rod-like chromosome ($X$) and a minute spherical mate ($Y$). In the case of macacus Painter found 48 chromosomes (Fig. 1, *a*); the male karyotype included, moreover, a pair of heteromorphous elements represented by a large rod-like $X$ and minute spherical $Y$, while the females possessed two isomorphous $X$ elements. Thus, Painter has established a close relationship between the karyotype of the monkey and that of man, both in number and morphology of sex chromosomes. However, he gave neither a detailed description of the separate components of the karyotypes of the monkey species, nor a comparison of the individual chromosomes with those of human karyotype.

Shiwago in turn also made a karyological analysis of the chromosomes in spermatogonial metaphases of the macacus, but, in contrast to Painter's data, he found the haploid number to consist of 42 elements. A detailed comparison of the 10 larger chromosomes with the corresponding elements of the human karyotype, carried out according to Andres-Navashin's scheme, has shown that, though some of these chromosomes are strikingly alike in their morphology to the point of being identical, the appearance of others differ considerably, which probably depends on the evolution of the karyotype of the primates. According to Shiwago, the morphology of the sex chromosomes is identical with that investigated by Painter.

Considering the meagre literary data bearing on this subject and divergency in opinions concerning some points of the question, we think it pertinent to publish our material comprising all the stages of spermatogenesis in macacus.

Testicles of two mature *Rhesus macacus** males, obtained from the Subtropical Branch of the All-Union Institute of Experimental Medicine at

* Determination of monkey species was carried out by Prof. M. F. Nesturkh.

Sukhumi, have been investigated. In one case the testicle was taken from an absolutely healthy animal; local novocaine anaesthesia was applied. In the second case—castration was combined with the amputation of one extremity; a few weeks previous to operation radium emanation was introduced into the cavity of the bone of this extremity, in order to induce a malignant neoplasm. The animal was operated under general ether anaesthesia. As it is the custom of this laboratory, the material was preserved with strong Flemming fluid containing less acetic acid than usually. Besides, we used one of the variants of Minushi's fluid containing: 2% osmic acid—4 parts, 1% chromic acid—8 parts and 3% potassium bichromate—8 parts. The material was passed into paraffin through chloroform. Sections 15 μ thick were stained with Heidenhain's iron haematoxylin, after Feulgen and with gentian violet after Newton. In both cases normal spermatogenesis was registered. 14 mitoses of the spermatogonia, 20 spermatocytes of 1st order and 8 spermatocytes of 2nd order were selected from a very large number of investigated preparations and mitotic figures; these were studied as to the chromosome number and chromosome morphology.

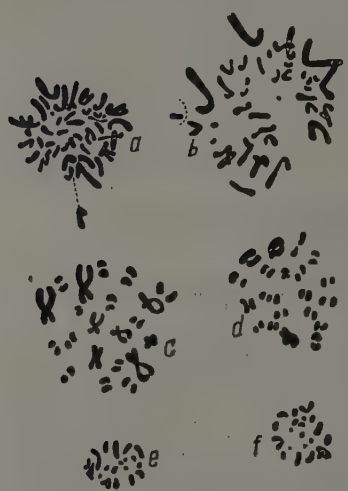


Fig. 1. *a*—Metaphase of spermatogonium (after Painter, Fig. 15), 48 chromosomes, including the heteromorphous pair. *b*—Metaphase of spermatogonium (original); 42 chromosomes, including heteromorphous pair. *c*—Metaphase of first spermatocyte (original); 21 tetrades. *d*—Metaphase of first spermatocyte (original); 21 tetrades; heteromorphous pair near the periphery of the plate. *e*—Metaphase of second spermatocyte (original); 21 chromosomes. *f*—Metaphase of second spermatocyte (original);22 chromosomes.

In most spermatogonial metaphases counted by us the number of chromosomes equalled 41—43 (Fig. 1, *b*). In three cases the chromosomes were more numerous and their number reached 45—50, and only in one polycentric mitose found in the animal which had been subjected to the action of the radium emanation, there were about 100 chromosomes present.

In most of first spermatocytes the number of tetrades was equal to 21 (Fig. 1, *c*); however, in some cells there was a surplus or a deficiency of 1—2 tetrades (Fig. 1, *d*). Finally, in the second spermatocytes, where accurate counts are difficult, the chromosomes being too small and crowded, we found 21—22 elements (Fig. 1, *e*, *f*). Lining the pairs of the chromosomes of spermatogonial equatorial plates according to their decrease in length, we succeeded, similar to Painter and Shiwago, in isolating a heteromorphous pair (Fig. 2). In our material the large *X*, having the form of a slightly curved rod or hook, often lay near the periphery, while the considerably smaller *Y* was usually found nearer to the centre of the plate. So far as the morphological characteristics of the separate components is concerned, our analysis coincides completely with that already published up by Shiwago. In the first spermatocytes the heteromorphous pair was usually found on the periphery of the metaphase plate and was distinguished by its shape and size; not infrequently the rod-like *X* was bent in the shape of a hoof, while the almost dot-like *Y* usually adhered to its mate. The above characteristics of

this tetrade and its specific disposition in the metaphase plate on its periphery is in good accordance with Minouchi and Ohta's (1939) data concerning the disposition of the said tetrade in human spermatocytes, which they considered typical of the heteromorphous pairs. Our observations show that the most probable orthoploid karyotype of the macacus is the 42 chromosome

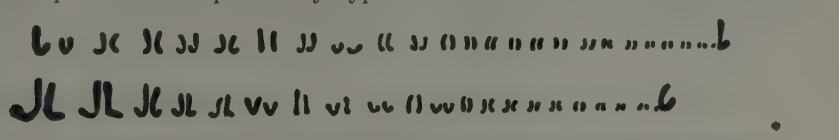

Fig. 2. Karyological analysis of metaphase shown in Fig 1, *a* (top); metaphase shown in Fig. 1, *b* (bottom).

set. The male possesses a pair of heteromorphous «sex chromosomes». Thus, our data are in complete agreement with those of Shiwago.

We are not inclined to ascribe any essential significance to fluctuations in the chromosome number involving 1—2 chromosomes. In the case of karyotypes such as those of primates, such deviations lie within the limits of possible errors; on the other hand, higher chromosome numbers suggest the presence of a certain number of heterosomic elements. Our material does not allow to judge of the causes responsible for such heterosomic elements, though their connexion with the non-disjunction of chromosomes is no less probable than in other cases of heterosomy. The hyperploid variant containing about 100 chromosomes might have been formed as a result of the abnormal course of mitosis. As it is known, polycentric mitoses with high chromosome numbers are not infrequent in the testes of homoiothermic animals. The fate of such cells in the spermatogenesis of man has been studied by Andres ([4]). Some of them lead to the formation of aberrant sperm cells, while others perish in the course of spermatogenesis.

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*PLANT PHYSIOLOGY*

# ENERGY EXPENDITURE BY FRUITS AS INDEPENDENT OF THE RATE OF RIPENING

By S. V. SOLDATENKOV

*(Communicated by A. N. Bach, Member of the Academy, 15. II. 1941)*

The ripening process as the final stage in the development of fruit is intimately connected with oxidizing respiration. In our former investigations ([1], [2]) we have found a definite quantitative relation between fruit ripening and the consumption of oxygen. The aim of the present work is to establish how the rate of ripening under various conditions of temperature and oxygen pressure is related to respiration as the unique source of energy.

The set-up used in the tests ([2]) made it possible to estimate the consumption of oxygen and the formation of carbon dioxide by fruits throughout the period of their ripening. The tomato fruits tested had been grown both in the greenhouse and in the open ground and carefully selected for similar outward appearance, size and weight. Tomatoes are a very convenient object for studying the ripening process because of their ability to grow ripe when stored. Their degree of ripeness can safely be estimated from their colour. But tomatoes will not usually get ripe all at a time, and so the fruits were removed singly as soon as they became ripe. This made possible to measure the consumption of $O_2$ and the formation of $CO_2$ at different stages of ripening.

Test 1. Balance of $O_2$ and $CO_2$ on fruit ripening in the air and in the medium enriched with oxygen under similar temperature conditions. The fruits of variety «Best of all» were sorted out into four lots, 27 in each, every lot ranging in weight from 1770 to 1790 g. Three of the lots were put into chambers of 4 l capacity. Two of these contained an atmosphere with 75% of oxygen; into the third air was admitted (21% of oxygen). One of the two chambers containing the oxygen-air mixture received ethylene in addition (1 : 2000). 250 $cm^3$ of a 4 *N* solution of potassium hydrate were poured into each of these chambers. The fourth portion was kept in the open (control). The chambers and the control fruits were put in a thermostat, $t° = 21—22°$. As the carbon dioxide was formed by the fruits, it was absorbed by the alkali solution, and the chamber, therefore, was automatically supplied with oxygen from a container, so that the concentration of this gas in the chambers was kept up at a constant level.

The results were as follows: with 75% of oxygen, 29.6% of the fruits were ripe on the 6th day; 92% on the 8th day. With 75% of oxygen plus ethylene 55% of the fruits had become ripe on the 6th day; 92% on the 8th day. In the chamber filled with the air (21% of oxygen) 37% of the fruits were ripe on the 8th day, 63% on the 10th day. Among the control fruits, which were kept in the air, 55% had become ripe on the 12th day only; 18% were yellow-pink, and the rest were green. The amounts of the oxygen consumed and of the carbon dioxide formed within the ripening period are

shown in Table 1. For all the three lots of fruits there is very close coincidence of data on oxygen consumption and carbon dioxide formation by 1 kg of fruits. If these data be compared with reference to the rate of ripening, it will be evident that the total amount of oxygen consumed and of carbon dioxide formed during the process of ripening are constant and independent of the rate of this process. The experiment was repeated, but the results were similar.

Table 1

| Ripening conditions | Gas consumed and formed in ripening in cm³ | | Gas per 1 kg in cm³ | | In % | |
|---|---|---|---|---|---|---|
| | $O_2$ | $CO_2$ | $O_2$ | $CO_2$ | $O_2$ | $CO_2$ |
| 75% of $O_2$ . . . . . | 9504 | 7876 | 5370 | 4449 | 110.9 | 105.8 |
| 75% of $O_2$ + ethylene (1 : 2000) . . . . . | 8387 | 7301 | 4700 | 4089 | 96.5 | 97.3 |
| 21% $O_2$ . . . . . | 8719 | 7523 | 4870 | 4201 | 100 | 100 |

Test 2. Balance of $O_2$ and $CO_2$ in fruits ripening at a different temperature: 24—25° (thermostat) and 12—13° (cellar). This time there were two lots, each including 16 fruits of the variety «Best of all». They had been grown in the greenhouse. Each lot ranged in weight from 1298 to 1295 g. They were kept in chambers. With regard to ripening the results were as follows: at 24—25° there were 56% of ripe fruits on the 8th day; 94% on the 9th day; with 12—13° there was 19% of ripe fruits on the 10th day; 37% on the 14th day; 63% on the 19th day; 94% on the 23rd day. In other words, the reduction of temperature reduced the rate of ripening of tomatoes.

The test was repeated with fruits grown in the open ground. Each of the two lots weighed 1430 g now. They were also kept in chambers. The $t°$ was 20° and 10—8° C. This time when the $t°$ was 20°, 20% of the fruits grew ripe on the 13th day; 74% on the 14th day; and on the 18th day all fruits were ripe. With 10—8° there was 33% of ripe fruits on the 28th day; 54% on the 35th day; 74% on the 38th day, and all fruits were ripe on the 45th day. In this experiment the dependence of the ripening rate upon temperature was still more conspicuous. When the temperature was 20°, the 74% ratio of ripe fruits was reached on the 14th day, while it was reached on the 38th day only, when the $t°$ was 10—8° C. Again, the rate of the ripening process was decreased more than 2 times owing to the reduction of temperature by 10—12°. The temperature coefficient of the ripening process conforms to the Vant-Hoff law.

The data on the consumption of oxygen and the formation of carbon dioxide within the ripening period in these two tests are presented in Table 2.

Table 2.

| Ripening conditions | Gas consumed and formed in ripening in cm³ | | Gas per 1 kg in cm³ | | In % | |
|---|---|---|---|---|---|---|
| | $O_2$ | $CO_2$ | $O_2$ | $CO_2$ | $O_2$ | $CO_2$ |
| 24—25° C . . . . . | 4245 | 3953 | 3271 | 3045 | 100 | 100 |
| 12—13° C . . . . . | 4380 | 4374 | 3382 | 3377 | 103.4 | 110.9 |
| 20° C . . . . . | 6290 | 6843 | 4398 | 4785 | 100 | 100 |
| 10—8° C . . . . . | 6517 | 6871 | 4557 | 4805 | 103.6 | 100.4 |

These data show that with different rate of ripening as determined by the temperature, the total amount of the oxygen consumed by and of the carbon dioxide formed by unit mass of the ripening fruit remains constant.

As the amount of carbon dioxide formed by the fruits is not affected by the rate of ripening, it is clear that the loss in organic matter by way of respiration (including the carbon dioxide) is not liable to be influenced by the rate of ripening, either. In fact, the estimations of losses in the dry weight of fruits on their ripening have shown that the same amounts of respiration material are consumed both at 20° and 10—8° C. The loss per 1 kg of fruits was 7.8 g in the first case, 7.7 g in the second. This gives a difference about 1.3%.

Whether the process of tomato ripening is carried on at different partial pressure of oxygen, or at different temperatures, the results obtained will reveal the same regularity: irrespective of the rate of ripening, the total balance between the consumption of oxygen and the formation of carbon dioxide remains constant.

The fact that the sum total of energy losses is independent of the ripening rate, can be expressed as follows: the product of respiration rate by the duration of ripening is a constant value: $R.\ T.$ =const. Here $R$ is the rate of respiration, expressed in terms of either oxygen consumption or formation of carbon dioxide; $T$ is the duration of ripening; the constant is dependent on the degree of development of the fruit. For fruits ripening under normal conditions, when the length of the period is about 1.5 or 2 weeks, this constant (in respect to either oxygen or carbon dioxide) is about 4.5 l per 1 kg of tomato fruits. It equals to about 3.4 l for fruits removed from the plant while near to the threshold of ripening, which get therefore ripe within a short time.

By increasing $R$ through the use of higher temperatures or higher partial pressures of oxygen, one can considerably reduce $T$, since the fruit will sooner use up the ration of oxygen required for ripening of its unit mass.

In the biological literature we find instances bearing analogy to the here established independence between the energy loss on ripening and the rate of this process ([3], [4]).

From the facts reported one can draw the conclusion that there is a very close relationship between respiration and the ripening process in fruits. The rate of ripening is influenced by the rate of respiration, and when the latter increases, the fruits grow ripe more rapidly. Therefore by controlling the rate of respiration in fruits with the aid of different factors (partial pressure of oxygen, temperature, etc.), one can give the process of ripening the desired direction.

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*AGROPHYSIOLOGY*

# INFLUENCE OF MANGANESE ON STRUCTURE OF YIELD IN SPRING WHEAT

By A. KALTIA

*(Communicated by A. N. Bach, Member of the Academy, 4. II. 1941)*

It has been established by experiments (1) that in order for plants to develop normally a number of minor elements, such as boron, manganese, zink, etc., are indispensable. The most thoroughly studied has been the influence of boron (2–5), which was found to affect not only the yield as a whole, but also the development of reproductive organs (6).

Our experiments were run with manganese. The plant tested was spring wheat *Lutescens* 062 grown in water and soil cultures.

In 1938 the tests were conducted in water cultures. The nutrient solution of Hellriegel was used, in which the dose of manganese (in terms of the pure element) was 0.05 mg per liter of the solution. Manganese was added in the form of sulphate. The object of the experiment was to trace the effect of manganese on the yield of spring wheat at different periods of development. The plan of the experiment and the results obtained are given in Table 1.

Table 1

| Lot | 1 | 2 | 3 | 4 | 5 | 6 | 7 |
|---|---|---|---|---|---|---|---|
| Manganese was available / Yield data | from the start until the outset of heading | from the start until the end of flowering | throughout the growing period | from the outset of heading till the end of flowering | from the end of flowering till the end of vegetation | from the start of heading till the end of flowering | without manganese (control) |
| Length of the growing period in days | 81 | 87 | 81 | 87 | 87 | 87 | 87 |
| Total yield, gs | 3.974 | 3.327 | 3.962 | 2.533 | 2.233 | 2.437 | 2.191 |
| Per plant in gs of air dry matter, % to the control | 181.4 | 151.8 | 180.8 | 115.5 | 101.9 | 111.2 | 100.0 |
| grain, gs | 1.353 | 0.815 | 1.181 | 0.633 | 0.485 | 0.488 | 0.373 |
| % to the control | 363.3 | 218.8 | 317.0 | 169.8 | 130.2 | 130.9 | 100.0 |
| Stem number per plant: total | 3.1 | 3.3 | 3.1 | 3.2 | 3.2 | 3.6 | 3.8 |
| spike-bearing | 3.0 | 3.1 | 3.1 | 2.9 | 3.0 | 3.4 | 3.6 |
| Spikelet number per spike | 13.9 | 12.8 | 13.4 | 10.9 | 10.2 | 11.1 | 10.2 |
| Grain number per spike: for all spikes | 18.6 | 11.9 | 14.9 | 8.8 | 6.5 | 6.1 | 5.0 |
| for grain filled spikes | 19.0 | 14.6 | 14.9 | 9.6 | 8.4 | 9.0 | 7.1 |
| Weight of 1 grain in mg | 24.31 | 22.27 | 25.62 | 24.73 | 24.66 | 23.61 | 20.88 |

As is seen from Table 1, the influence of manganese on the plant as a whole and on the development of its reproductive organs was favourable, especially if it had been given at the beginning of vegetation. Though the plants responded to introduction of manganese with reduced tillering (as compared to the controls), and the first lot of plants had an average of 3.1 stems per plant, as against 3.8 stems in the lot 7, nevertheless the total yield of plants increased considerably. As for the structure of the spike, the effect of manganese introduced at the beginning of vegetation was well expressed, too: the number of spikelets was higher and so was the number of grains in a spike (lot 1). The influence of manganese was weaker if this element had been introduced at later dates, when the spikelets had already been formed; but even then the number of grains showed an unmistakable increase (lots 4—7). The presence of manganese in the nutrient solution did not only increase the number of grains per spike, but also improved their quality by raising their absolute weight.

In 1940 the tests were run in soil cultures (Table 2).

Table 2

| Yield data \ Lots | In terms of the pure element | | |
|---|---|---|---|
| | Minus manganese | 4 mg of manganese per 1 kg of soil | 2 mg of manganese per 1 kg of soil |
| Stem number per pot (10 plants): | | | |
| total | 32 | 33.5 | 34.5 |
| spike-bearing | 23.7 | 27.5 | 28.7 |
| Spikelet number per spike | 16.1 | 15.8 | 15.5 |
| Grain number per spike | 12.7 | 13.5 | 12.3 |
| Weight of 1000 grains in g | 36.25 | 34.82 | 36.40 |
| Weight of air dry matter per pot (10 plants): | | | |
| grain | 10.91 | 12.93 | 12.88 |
| spikes | 3.68 | 4.35 | 4.71 |
| stems and leaves | 32.71 | 37.31 | 38.91 |
| total | 47.24 | 54.59 | 56.26 |
| % to the control: | | | |
| grain | 100.00 | 118.51 | 118.05 |
| total weight | 100.00 | 115.55 | 119.10 |

From these data it is obvious that under soil conditions the effect of manganese is not so strong as in water cultures, though it still remains favourable. If in the case of water cultures the effect is due to a better filling of the spike, in this case it is the number of spike-bearing stems which makes for a higher yield. As before, we found that the plus-manganese plants were by far less strongly affected with mildew, though the difference was not so conspicuous. The conclusion to be drawn from these experiments is that the development of reproductive organs in spring wheats is promoted by manganese; there is an increase in the absolute weight of grain, and the green mass of every stem is better developed. This goes to increase the total yield and to reduce the affection of the plant by mildew.

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*PHYTOPATHOLOGY*

# BIOCHEMICAL CHANGES IN THE CASE OF ANTHOLYSIS

By V. L. RISCHKOV and E. P. GROMYKO

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In plants affected with virus diseases the phenomena of antholysis and proliferation are often encountered. These transformations, as shown by Rischkov, occur mainly in such virus diseases, in which there is observed affection of floem, and carbohydrates are detained in the leaves. The author has expressed the assumption that the reason for antholysis in such cases is the poor supply of carbohydrates to the developing inflorescence, whereby the normal ratio of C/N is disturbed (by the ratio C/N we mean the relation between the sum of carbohydrates and the total nitrogen). In order to verify this assumption, we have analyzed the organs of oats affected with the virus disease of zakuklivanie, in which the proliferation of flowers is known to be attended by a reduced outflow of carbohydrates from the leaves (according to V. L. Rischkov, M. N. Vorobieva and E. P. Gromyko).

The experimental material consisted of dry leaves from affected and sound oat plants, as well as of spikelets. Leaf sampling was done at two different times: the first time at the period of tillering, the second time at the period of stalk formation. The spikelets were collected before they came out of the tube formed by the leaf sheeth lest the assimilation of carbohydrates by them should take place and thereby obscure the picture of their supply with carbohydrates by the green parts of the plant. The carbohydrates were estimated according to A. R. Kiesel; the total nitrogen according to Moore, and the protein nitrogen according to Barnstein.

In Tables 1 and 2 are given the data on the content of carbohydrates and nitrogen in the leaves of sound and affected plants, while in the phase of tillering. From these data it is evident—what might have been expected, indeed—that the leaves of affected plants are much higher in carbohydrates and lower in nitrogen than are the leaves of sound plants. At the time of stalk shooting both the sound and affected leaves increase in carbohydrates and decrease in nitrogen, the change being more pronounced in the plants affected. As for the spikelets, there are more heavily stored with carbohydrates when in the sound condition than if affected with zakuklivanie, whereas the amount of total and protein nitrogen is nearly the same in both cases.

It is obvious that in the affected plants the inflorescence is insufficiently supplied with carbohydrates, owing probably to a severe affection of the phloem and the inhibited outflow of carbohydrates from the leaves, which are typical symptoms of this disease.

Table 1

Content of Carbohydrates in Sound Oats and in Oats Affected with Zakuklivanie (% by weight in dry matter)

| Material | Reducing sugars | Saccharose | Starch | Total |
|---|---|---|---|---|
| Tillering, sound leaves | 1.25 | 8.54 | 0.02 | 9.81 |
| Tillering, affected leaves | 2.07 | 10.67 | 0.42 | 13.17 |
| Stalk formation, sound leaves | 4.35 | 7.60 | 0.73 | 12.68 |
| Stalk formation, affected leaves | 8.46 | 12.03 | 2.85 | 23.34 |
| Stalk formation, sound spikelets | 13.95 | 19.98 | 2.67 | 36.40 |
| Stalk formation, affected spikelets | 8.48 | 3.95 | 3.14 | 15.50 |

Table 2

Content of Nitrogen (% by dry weight) and C/N Ratio in Sound and Affected Oat Plants

| Material | Total nitrogen | Protein nitrogen | C/N |
|---|---|---|---|
| Tillering, sound leaves | 4.94 | 4.53 | 1.9 |
| Tillering, affected leaves | 3.79 | 2.95 | 3.4 |
| Stalk formation, sound leaves | 4.44 | 3.92 | 2.85 |
| Stalk formation, affected leaves | 3.17 | 2.46 | 7.50 |
| Stalk formation, sound spikelets | 2.62 | 2.03 | 13.95 |
| Stalk formation, affected spikelets | 2.99 | 2.23 | 5.20 |

These data seem to confirm the theory of Klebs as to the importance of C/N in flowering. Many authors, as Knodel, Polster and others, have failed of late in their attempts to find evidence in support of Klebs' theory by estimating the C/N ratio in the leaves of various plants at the time of flower initiation. We are by no means inclined to consider the C/N ratio as the unique factor responsible for the transition to flowering. Nor do we believe that it may play a more important part or be endowed with a more specific influence on the transition of plants to the reproductive phase, compared to a variety of other conditions. It appears, however, that no normal development of flowers is possible if the ratio between carbohydrates and nitrogen is disturbed. It is the disturbance of the C/N ratio that accounts for antholyses, which so often attend virus diseases of the yellow type. In the case of these diseases carbohydrates are arrested in the leaves, and the developing inflorescence is not supplied with sufficient amounts of carbohydrates, when they are most needed by it. Its development proceeds in an atypical way and leads to various forms of flower antholysis and of proliferation.

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*GENETICS*

# CHROMOSOMAL ABERRATIONS, LETHAL MUTATIONS AND X-RAY DOSAGE

By N. P. DUBININ, V. V. KHVOSTOVA and V. V. MANSUROVA

*(Communicated by I. I. Schmalhausen, Member of the Academy, 15. II. 1941)*

The present paper is concerned with the frequency of chromosomal aberrations and of lethal mutations (in the X-chromosome of *Drosophila melanogaster*), as dependent on the X-ray dosage. The purpose of the present work was to obtain complex material on this problem. There has been studied the effect of six different dosages of X-rays, viz. 250, 500, 1000, 2000, 4000 and 6000 r, both on the frequency of chromosomal aberrations and on that of sex-linked lethal mutations.

Males of the inbred line *forked* crossed with $\frac{ClB}{sc^{8}w}$ females were X-rayed. The $F_1$ larvae were examined for the presence of chromosomal aberrations by analyzing the structure of the salivary gland chromosomes on provisional acetocarmin preparations. Lethal mutations were detected in $F_1$ by the *ClB* method. Thus, regularity of the increase in the number of chromosomal aberrations and of lethals, according to the increase in dosage, was studied on the same material. About 100 lethals from each dosage were examined for chromosomal aberrations (microaberrations were practically not taken into account when the $F_1$ larvae and lethals were examined).

A total of 4137 larvae were examined and 425 structural changes were found connected with 917 chromosome breaks (the latter were located according to Bridges' map). The results obtained show clearly that there is no direct proportionality between the quantitative increase in the number of chromosomal mutations and that of the X-ray dosage.

29481 $F_1$ cultures were examined for lethals and 1270 lethal mutations were found. The number of lethal mutations increases in direct proportionality to the X-ray dosage; if we take into account the induction of double lethals, we obtain a picture of an ideal direct proportionality. In examining the structure of lethal bearing chromosomes, we have found that lethals induced by different dosages of X-rays show various relation to chromosomal aberrations: the stronger the dosage, the greater is the number of lethal mutations connected with chromosomal aberrations (Fig. 1).

The analysis of distribution of breaks in chromosomal aberrations has shown that the number of breaks per chromosome is in strict direct relation to its length, this relation being particularly well pronounced if the chromosome length is measured on salivary glands.

The great number of breaks, corresponding to the inert regions within a chromosome, is in disproportion to its relative length. A certain concentration of breaks is also observed in the distal regions. A detailed analysis of the distribution of breaks within the regions of Bridges' map shows microsections with a breakage frequency disproportionate to the length of the

section (for the *X*-chromosome this phenomenon has been described by Prokofieva-Belgovskaya and Khvostova and also by Kaufmann). Similar breakage points have been found in chromosome II: 21D, 22A, 22B, 22D, 24F, 57F and 59D; in chromosome III: 61F, 62B, 64B, 61C, 61D, 99F, 99AB, 93E—F, 89D—E.

Kaufmann has proved that the corresponding microsections in the *X*-chromosomes are apparently connected with duplications. Similar conditions may exist in other chromosomes, as well. An analysis of the plural breakage distribution within the nucleus shows that they do not arise by chance. It is evident that the breaks are concentrated within one chromosome. In the case of several breaks taking place within the nucleus, they mostly produce no complex common aberration, but rather form separate independent aberrations of a comparatively simple structure. This circumstance points to the existence of a certain spatial limitation in the process of formation of aberrations.

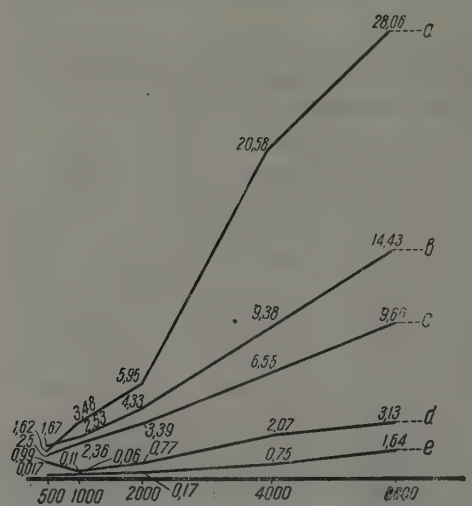


Fig. 1. Curves showing the result of analysis of the relation between the number of chromosomal aberrations and lethals and X-ray dosage. Top line (*a*) shows the disproportionally high increase in the number of chromosomal aberrations; second line (*b*)—the experimental increase in the per cent of lethal mutations; third line (*c*)—the dependence of lethals free from chromosome aberrations on the dosage; fourth line (*d*)—the increase of the per cent of lethals causally connected with aberrations; fifth line (*e*)—the per cent of cases when a chance coincidence of two lethals is observed.

The rate of chromosome mutation shows departures from the direct proportionality to the X-ray dosage, while between the frequency of lethal mutations and the dosage there is an ideal linear relation. The data of the present paper show, moreover, that a considerable group of lethal mutations is connected with structural changes in the chromosomes.

The most reasonable interpretations of the variation in the number of chromosomal aberrations, on the one hand, and in that of lethal mutations on the other, according to X-ray dosage, is to assume that these processes express a different effect induced by X-rays on the chromosomes.

It has been shown in a number of papers that the rate of visible and lethal mutations is directly proportional to the dosage expressed in ionisation units (r), and is not affected by the wave length; hence one may infer that X-ray induced mutations are due to ionisation or atom excitation only.

Sax has shown on *Tradescantia* the number of single breaks to be also in direct proportion to dosage. Eberhardt has established that the number of breaks resulting in the position-effect of the gene for cubitus interruptus does not depend upon wave length, being in direct proportion to the degree of ionisation only.

Despite the fact that the group of lethal mutations proves to include different categories of changes, it is quite possible that it is induced by the

primary X-ray effect connected with single «hits». This is quite evident, in so far as gene mutations are concerned. Small deficiencies and microaberrations can be induced by a single «hit», touching the ring of a small spiral. As to lethals connected with chromosomal aberrations, it should be taken into account that the fragments produced by a single «hit» may reconstruct the initial structure, or give an aberration, if there is another break in the nucleus. However, a lethal change may take place in both cases; thus, in the course of a process, where the secondary effect in the form of joining fragments leads to chromosomal mutations, the occurrence of lethals may be due to the primary effect of X-rays.

The complicated relation existing between the number of chromosomal aberrations and the X-ray dosage as considered by us, supports Stadler's hypothesis (1932) of the fragmentary nature of chromosomal aberrations. This hypothesis has been experimentally worked out on *Tradescantia* by K. Sax (1938, 1940), who has shown the rate of chromosome mutation to be proportional to the square of the dosage, while the number of primary breaks is directly proportional to the X-ray dosage.

In our case there is also a clear distinction as to the dosage between mutations induced by the primary effect of X-rays (lethal mutations), and chromosomal aberrations connected with the secondary effect of irradiation (recombination of fragments). The fact that chromosomal aberrations in *Drosophila melanogaster* show no proportionality to the square of the dosage is quite natural, since the combination of fragments may become complicated by a number of intracellular conditions.

It would be natural to explain the data obtained by assuming the fragmentary nature of the process of formation of chromosomal aberrations and regarding lethal mutations as being induced by the primary effect of X-raying. In order to elucidate the nature of the formation of chromosome mutations, it is also important to make clear the rôle played by the time factor in the appearance of chromosomal aberrations.

In full accord with the fragmentation theory, Sax has found that in the case of *Tradescantia*, the time factor does not affect the frequency of single breaks, while it does affect the number of chromosomal aberrations. However, the experiments with *Drosophila* have not yet given such clear results. The data obtained by Timofeeff-Ressovsky (1936) show that the time factor plays no part in the formation of lethal mutations. However, he studied lethals induced by dosages of 3000—3600 r, about $\frac{1}{4}$ of which must have been connected with chromosomal aberrations. Our preliminary data obtained from a study of $F_1$ larvae, while the dosages were biologically controlled, i e. the percentage of lethals was analyzed, also testify to the absence of the effect of the time factor upon the induction of chromosomal aberrations. Eberhardt's data (1939) do not agree with the above-mentioned. However, his experiments involved no biological control of the dosage.

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*GENETICS*

# STUDY ON THE NATURE OF CHANGES CAUSED BY STRUCTURAL CHROMOSOME MUTATIONS

By B. N. SIDOROV

*(Communicated by I. I. Schmalhausen, Member of the Academy, 15. II. 1941)*

The most common type of changes caused by the position effect is the decrease in dominance of the normal allelomorph. One of the cases of the position effect, which has been most thoroughly investigated, viz. the position effect of the gene for cubitus interruptus (*c. i.*), belongs to this type of changes. It may be assumed that the lack of data on the position effect causing an increase in the dominance of the normal allelomorph, is accounted for by the difficulties encountered in disclosing such cases. These difficulties are connected with the fact that, as a rule, normal allelomorphs are completely dominant.

The work described in the present paper is an attempt to verify this assumption on a case of the position effect of the gene for cubitus interruptus. With this aim in view I arranged to obtain changes producing the position effect of the recessive allelomorph of *c. i.* As dominant changes of the *c. i.* type ($c.\ i.^D$) are practically known to exist, it seemed quite natural to expect that in case the action of the recessive *c. i.* allelomorph is enhanced by the position effect, it will easily be detected in compound with a normal allelomorph.

In order to disclose such changes in the recessive *c. i.* gene (transition from recessive to dominant state) *ey ci* males were X-rayed (4000 r) and crossed with normal females. The experiment went fully to confirm our anticipation. Among 26612 $F_1$ flies 72 (0.25%) were to a varying degree endowed with the character *c. i.* A genetic analysis of these aberrant flies has shown them actually to possess the dominant changes of *c. i.* Part of this material was analyzed both genetically and cytologically on the salivary gland chromosomes. In 5 cases out of the 18 examined no changes of chromosome structure could be revealed, neither genetically nor cytologically. These changes are apparently point mutations of the *c. i.* gene. In the rest of the cases there were found chromosomal aberrations, connecting chromosome IV with another chromosome. In these aberrations the break in chromosome IV occurred near the *c. i.* gene, in the same point as in the cases of the position effect of the normal allelomorph of this gene. Another break occurred in all the cases studied in the active regions of other chromosomes. Thus, an increase in dominance of the recessive *c. i.* is caused by the same type of aberrations which produce the position effect of the usual type (decrease in dominance) in the normal allelomorph.

Further on the production of dominant changes of the *c. i.* gene made it possible to disclose changes of the normal allelomorph connected with structural changes in the chromosomes and tending to an increase in dominance.

It is known ([3],[5],[6]) that aberrations possessing a break in chromosome IV near the gene for *c. i.* do not give the usual position effect, if another break had taken place in the inert region, or even in the active region adjoining the inert one. As test object I took two such chromosomal aberrations between chromosomes III and IV (No. 1049—break in chromosome III according to Bridges' map in 84A, and No. 81—break in the inert region) and studied the manifestation of the *c. i.* character in compound with three dominant changes of this gene. The results are given in Table 1 (with translocation No. 1049) and Table 2 (No. 81). Compounds of these translocations with dominant changes of *c. i.* and with a normal chromosome were produced in the same cross.

Table 1

| Structure of the flies | Manifestation of *c. i.* in females, in % | Manifestation of *c. i.* in males, in % |
|---|---|---|
| Tr. 1049 / *c. i.* 65 | 11.4 | 22.1 |
| + / *c. i.* 65 | 55.6 | 77.5 |
| Tr. 1049 / *c. i.* 25 | — | 18.5 |
| + / *c. i.* 25 | — | 49.2 |
| Tr. 1049 / *c. i.* 45 | 0.0 | 2.7 |
| + / *c. i.* 45 | 6.0 | 40.3 |

Table 2

| Structure of the flies | Manifestation of *c. i.* in females, in % | Manifestation of *c. i.* in males, in % |
|---|---|---|
| Tr. 81 / *c. i.* 65 | 3,2 | 7.3 |
| + / *c. i.* 65 | 19.1 | 38.5 |
| Tr. 81 / *c. i.* 25 | — | 10.4 |
| + / *c. i.* 25 | — | 30.0 |
| Tr. 81 / *c. i.* | 0.0 | 0.76 |
| + / *c. i.* 45 | 18 | 18.7 |

Tables 1 and 2 show that the manifestation of the *c. i.* gene is strongly suppressed in the compounds of translocations with any dominant changes of *c. i.* as compared with structures heterozygous for the dominant *c. i.* and the normal chromosome. Thus, aberrations of this type cause an increase in dominance of the normal allelomorph. The discovery of this new type of position effect of the normal allelomorph shows that any break near the *c. i.* gene practically causes some change in its action. It even seems to me possible to affirm that in a large number of cases the existence of the position effects cannot be detected by the usual methods, and that as a matter of fact the phenomenon of the position effect is of much more frequent occurrence than it has been supposed heretofore.

Further studies of the position effect of the *c. i.* gene lead to the solution of another important question; namely, we succeeded in finding out that the changes in the *c. i.* gene, due to changes in its position, differ markedly from the point mutations of this gene. It has already been shown ([2], [4]) that both the dominant (*c. i.*$^{D}$) and the recessive mutation (*c. i.*) are qualitative changes (neomorphous, according to Muller's terminology). The absence of chromosomal aberrations in these mutations has been shown by Bridges ([1]). I undertook a special investigation in order to clear the character of the changes of the normal allelomorph of the *c. i.* gene connected with chan-

ges in its position. In case these changes would also prove to be of neomorphous character, as the *c. i.* and *c. i.*$^{D}$ mutations, the manifestation of the *c. i.* character may be increased by the accumulation of normal allelomorphs of this gene, changed under the influence of the position effect. If the manifestation of *c. i.* will, on the contrary, decrease as the result of this accumulation, one may conclude as to the purely quantitative change of the normal allelomorph (hypomorphous changes, according to Muller). I obtained flies containing one chromosome IV with a *c. i.* mutation, and two normal allelomorphs changed by the position effect, from translocations No. 37 (II—III) and No. 684 (III—IV). The manifestation of *c. i.* in this structure was much feebler than its manifestation in compound with each of these translocations separately. Comparative results are given in Table 3.

Table 3

| Structure of the flies | Frequency of occurrence in % | | | | | |
|---|---|---|---|---|---|---|
| | In females | | | In males | | |
| | first region cubitus | third region cubitus | analis | first region cubitus | third region cubitus | analis |
| $\frac{\text{Tr. 684}}{c.\ i.} = 200$ | 61.5 | 97.5 | 70 | 70.0 | 100 | 75.5 |
| $\frac{\text{Tr. 37}}{c.\ i.} = 200$ | 33.0 | 71.0 | 10 | 47.5 | 75 | 21.0 |
| $\frac{\frac{\text{Tr. 37}}{\text{Tr. 684}}}{c.\ i.} = 200$ | 23.5 | 50.5 | 0.0 | 28.5 | 57 | 13.5 |

Table 3 gives the per cent of changes of the cubital and anal veins. For the cubital vein the percentage is given for two regions separately. The first region extends from the base of the wing to the first transversal, and the third region—from the second transversal to the end of the wing. This experiment conclusively proves the difference between the changes of the normal allelomorph of the *c. i.* gene caused by the position effect, and the point mutations of this locus.

The discovery of the existence of such a difference between the character of the changes connected with structural chromosome mutations and point mutations, seems to me to disprove the mechanistic conception of Goldschmidt and some other investigators, who try to reduce hereditary variation to mere structural changes of the chromosomes.

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*GENETICS*

# ON THE LAW OF EXPANSION AND EXTINCTION OF GENIC ACTION

By J. A. RAPOPORT

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The specificity of the effect of the genes consists in that the action of each mutation affects mainly a definite trait of the living organism. Very often the effect of a mutation is restricted to a change in some peculiarity of a given organ or of a part of the body. The locality, spatial limitation of the action of the gene may be explained in different ways. It may be due to the purely topographic character of the function of the gene, viz. to its influence on a certain, always the same, spatial region of the embryo. On the other hand, this locality may be explained by the qualitative character of the gene's action, by its capacity to affect only those substances or structures which underwent a peculiar determination and show a definite chemical composition, almost irrespective of their site in the embryo. The problem whether there exists some relation or concordance in the action of the genes which affect the development of a given organ, as opposed to the relation between genes affecting different parts of the body, has not been analyzed. Do they form a special complex and a system in the process of development, or not? The qualitative mode of action of hereditary factors makes the existence of such systems possible. In the case of a positive solution of this problem, besides the linear grouping of genes in a chromosome, which formed a link between genetics and cytology, another possibility, that of establishing functional developmental groupings, will be available. Such a functional grouping, in the case of correspondence between organs and these phenogenetic groups, will certainly make students in comparative anatomy feel more interested in genetic phenomena, than they have been hitherto.

For an experimental study of this problem new methods are required. We used the phenogenetic method of analysis. The presumptive rudiment of one organ was experimentally transformed into another, and on the transformed material of this rudiment there was studied the effect of mutations, specific both for the organ into which the rudiment would have developed during the usual course of the process, and for the part into which it has been experimentally transformed. Homeioses, for instance aristapedia, which were formerly utilized by us with the same purpose, are unsuitable here, since they are atavistic changes. For the rudiments of organs affected by homeiosis there exist genotypically two developmental systems—a modern

one and a more ancient one which has not disintegrated still. As test object we used the dominant mutation *Met* (chromosome II), lethal under homozygous condition, which we discovered several years ago. Its most characteristic feature is the transformation of the presumptive material of the wing into the mesonotum. The wing is absent in a large number of *Met* individuals, while the mesonotum is transformed into a gigantic formation which absorbs the potential material of the wing. The extra mesonotal structures are covered with macrochetae and microchetae, and in the details of their structure they do not differ from the material of the normal back. A detailed analysis of this mutation (along with some others) has shown that the embryogenesis of the representatives of the order *Diptera* which have so far been considered as a model of the mosaic type of development, are characterized by the same phenomena of induction, regulation, etc., which are known for the regulatory types.

Combinations of *Met* with *Bx*, *m*, *va*, *cv*, *dir*, and twenty more, including from 80 to 6000 individuals per combination, caused no changes in the structural peculiarity of the giant mesonotum. In all the cases studied, the mutations which, as a rule, cause pronounced and obvious changes of various wing characters, show no action on the structure of the giant mesonotum, after the presumptive substance of the wing is transferred into another morphogenetic system. The effect of shape and size mutations, of dominants and recessives, is soon eliminated, as well. Thus, the transformation of the potential material of organ *A* into that of *B* extinguishes the action of the genes which determine the specific peculiarities of *A*. Experiments carried out in another direction are of interest, too. These experiments were arranged to elucidate the mode of influence of the mutations specific for organ *B* on that part of the structure which was formed out of the presumptive material of organ *A*. For these experiments we have chosen mutations, or allelomorphs of mutations which affect the shape and size of the dorsal bristles, without changing the structure of the wing. Combinations of *Met* with *f*, *ss*, *M*, *Sb*, *spl* and many other mutations were studied, including up to 8000 individuals in some combinations. Three to five extramacrochetae on the giant mesonotum had the structure of their bristles corresponding to *f*, *ss*, etc. The following conclusions seem justified: During the transition of the *A* material into the embryonic complex of *B*, the specific mutations of the *B* organ automatically expand their action on the included substance. An expansion of the influence of some genes beyond the usual limits takes place.

Thus, the experiments show that there exist definite complexes and systems of mutations whose members are functionally and causally connected with one another, since they cannot be singly excluded from or included into a process. Unlike the system of genes within a chromosome, which affect but separate traits of the organism, these genetic systems affect whole parts and organs of the organism, constituting the traditional object of morphology and physiology. A series of new linkages and connexions is revealed among genes. But this series is no morphological one, and it finds no expression in the chromosome map; it is dynamic and corresponds to the genes in action. The mutations affect the forming structures according to the qualitative, determinative specificity of the rudiment, rather than to the topographic principle. If a volume of the material, twice or thrice as large as the normal, expanding far beyond the limits of the normal rudiment or even located in another place, is determined as a definite rudiment, then all the peculiarities of this rudiment will develop in it. On the other hand, irrespective of the reduction of the volume of the determined rudiment, mutations specific of the morphogenic system of the rudiment, will become

manifested in it. Thus, the remaining, sometimes minute regions of the wing tissue may be *dir*, *m*, or *Bx*, etc., according to the genotype.

It is most essential that the fate of single recent mutations, utilized in genetical experiments, and that of phylogenetic combinations of hereditary changes which cause the normal development of a given organ, do not differ in the transition of the presumptive substance from one morphogenic unit to another. The transition is equally radical for each of them, and their action is equally extinguished. This concerns also the expansion of the gene effect in the course of such transitions. The similar behaviour of the historical (evolutionary) and experimental (genetic) differences may be regarded as an important proof of the assumption of their common nature and of evolutionary changes, being formed out of mutations by natural selection.

We believe that the above-mentioned facts, besides their independent phenogenetic importance, are of interest from the point of view of some concepts of comparative anatomy. This may be illustrated on two problems: substitution and convergence. The transformation of the wing material into the mesonotal one may be regarded as one of the ways of substitution. For comparative anatomy it is of interest that here the transition is at the same time a morphological and a functional one. During substitution one organ disappears or becomes greatly reduced, while another one grows and a whole system of genes, affecting the structure of the organ as a whole, at once manifest their action in the new domain. Certainly, this does not exclude other modi of evolutionary changes of organs adopted by Dohrn ([1]), Kleinenberg ([2]) and others.

Thus, genetics is capable to prove experimentally the reality of the process of substitution, and to reveal new regularities of the phenomena which have been disclosed by comparative anatomy through deduction. Moreover, by methods of its own, it can discover the connexions existing in the process of formation of highly integral complexes—parts and organs—which are the object of morphology.

Quite puzzling is the appearance of great mutational leaps, the so-called macromutations, omnipotent mutations etc., which sharply alter the structure of an organ or of a whole system, and may become adaptative in a favourable medium. According to some authors, e. g. Goldschmidt ([3]), it is only of such changes that the progressive evolution was built up. Without denying the value of such changes which may occasionally take the place of idioadaptations or even aromorphoses, we believe these shifts to be of secondary nature. A primary great adaptive change is also formed by leaps, but gradually, including a number of leaps. For instance, for the development of a completely new organ out of an old one, according to Dohrn's principle, or out of a newly arisen rudiment, a considerable number of mutations are necessary, which must model and perfect all the details of the structure. This is the only thinkable way for a completely new organ to form. Different ways may be taken if a system formed in such a way were to disappear in the course of speciation, having become noxious or superficial in a new environment, as the wing disappears in *Met*, or in a more complicated way. The reappearance of the effaced structure can take place much more rapidly than its original development. It may become fixed under a somewhat different aspect by a small number of leaps. Therefore, adaptive and complicated macromutations, resulting in a sudden formation of new organs, are secondary shifts, unlike the primary process, but utilizing those mutations continuing to function in hypostatic condition and numerous phenogenetic connexions formed historically in the course of the primary process. Consequently, though the same structure of an organ is not repeated in all its details in the phylogenesis, convergence and parallelism are

not merely external similarities, but are frequently founded on common or successive hereditary and developmental systems, or even on separate identical elements of these systems. Convergence is often grounded on a common genetic and phenogenetic basis. Thus, although, according to the data of segregation, the action of genes is scattered at random throughout various characters of the organism, in the process of development these separate reactions are united into groups which correspond to organs and systems of the body, and are mediated by the development of the latter.

And though the action of various genes is interwoven and overlapped, some genes being important for the development not of one, but of several structures, the particular systems of genes described in the present paper behave as a whole in the course of different transitions, this being important not only for ontogenesis, but for phyletic changes, as well. The phenomena of extinction and expansion of the action of genes, as described here, are a function of the existence of separate phenogenetic systems and chains which can be discovered by their help.

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*ZOOLOGY*

# VERTICAL DISTRIBUTION OF GASTROPODS IN THE SOUTHERN PART OF THE CASPIAN SEA

By V. P. KOLESNIKOV

*(Communicated by A. A. Borissiak, Member of the Academy, 15.I. 1941)*

A rich collection of molluscs, chiefly gastropods (26695 specimens), was gathered by A. F. Nossov in 1933, during the dredgings of the s/s «Sextant». The shells were collected on 59 stations, of which nineteen were situated along the western coast of the Caspian Sea, between the strait of Apsheron and Lenkoran, at depths ranging from 4 to 40 m (17987 specimens, with a mean of 946 per station); six between Baku and Krasnovodsk, depth 40—200 m (1149 specimens, mean 195); five between Lenkoran and Ogurchinsk island, depth 40—200 m (684 specimens, mean 134); thirteen between the same point but at depths of 200—975 m (16 specimens, mean 1) and sixteen on the eastern coast of the Caspian Sea between Krasnovodsk peninsula and Ogurchinsk island, depth 5—40 m (6859 specimens, mean 429).

The greatest quantities of gastropods have been found on the western coast (55.6% of the number of an average station), a far lesser number on the eastern coast (25%) and a relatively very small number between Baku and Krasnovodsk (11.5%), and between Lenkoran and Ogurchinsk island (7.9% at 200 m). More than 80% of gastropods were found at depths ranging from 4 to 40 m, and only 20% at 40—200 m.

There are some peculiarities about the vertical distribution of gastropods, which as yet refuse proper explanation. Thus, for instance, there is a sharp decrease in the number of shells below 20 m (718 specimens at 20—40 m), with a minimum at 60—80 m (54 specimens), while beyond 80 m their number rises again, reaching a new maximum at 150—200 m (821 specimens). Only occasional specimens are found at depths greater than 200 m. However, this value of 200 m must be regarded as approximative, since there is a gap in our observations between 200 and 347 m, and the lowest limit of distribution of gastropods cannot be exactly determined. It must be assumed to lie between 200 and 347 m.

The greatest quantity of shells corresponds to depth of 10—20 m (17387); above 10 m their number lessens considerably (6735), while below 20 m there is a sharp drop of the number of representatives of *Theodoxus*, *Caspia*, *Lithoglyphus* and, especially, *Hydrobia*. However, the generic composition remains unaltered, and only beginning with 40 m the character of the fauna undergoes some changes, viz. *Clessiniola martensi* (Dyb.), *C. variabilis* (Eichw.), *Turricaspia sieversi* (Cless.), *T. pseudodimidiata* Dyb. et Gr. disap-

pear or become very scarce, and new species appear: *T. bakuana* n. sp. and *T. dimidiata* (Eichw.). Still deeper (60—200 m) there is a complete change of the character of the fauna: *Theodoxus pallasi* Lindn., *T. (Ninnia) schultzi* (Grimm), *Hydrobia ventrosa* Montg., *H. grimmi* Cless., *H. chrysopsis* n. sp., *Turricaspia bakuana* n. sp. disappear along with the species mentioned above; *Lithoglyphus exiguus* (Eichw.) are very rare, but *Clessiniola grimmi* (Dyb.), *Clathrocaspia pallasi* (Dyb.), *Turricaspia grimmi* (Dyb.), *T. dimidiata* (Eichw.), *Planorbis* (*Gyraulus*) *eichwaldi* Grimm occur in relatively great numbers.

Vertical Distribution of Species with Reference to the Number of Specimens

| Species | Depth in m | | | | Total |
|---|---|---|---|---|---|
| | 4—20 | 20—80 | 80—200 | 200—975 | |
| *Theodoxus pallasi* Lindh. | 800 | 23 | — | — | 823 |
| *T. pallasi* Lindh. var. *nalivkini* n. var. | 11 | 5 | — | — | 16 |
| *T. (Ninnia) schultzi* (Grimm) | 77 | 22 | — | — | 99 |
| *T. (N.) schultzi* (Grimm) var. *jukovi* n. var. | 2 | — | — | — | 2 |
| *Hydrobia venirosa* Montg. var. *evanescens* nom. nov. | 412 | — | — | — | 412 |
| *H. grimmi* Cless. | 6375 | 55 | — | — | 6430 |
| *H. grimmi* Cless. var. *gemmata* nom. nov. | 8999 | 92 | — | 4 | 9095 |
| *H. chrysopsis* n. sp. | 472 | 4 | — | — | 476 |
| *Lithoglyphus exiguus* (Eichw.) | 1505 | 32 | 2 | 2 | 1541 |
| *Zagrabica brusiniana* Dyb. | 616 | 180 | 487 | — | 1283 |
| *Clessiniola mariensi* (Dyb.) | 18 | — | — | — | 18 |
| *C. variabilis* (Eichw.) | 7 | — | — | — | 7 |
| *C. grimmi* (Dyb.) | 18 | 5 | 118 | — | 141 |
| *Caspia caspia* (Eichw.) | 39 | 14 | 6 | — | 59 |
| *C. inflata* (Dyb. et Gr.) | 131 | 4 | 41 | — | 176 |
| *C. conus* (Eichw) | 3445 | 312 | 93 | 2 | 3852 |
| *C. cincta* (Abich) | 394 | 23 | 29 | 2 | 448 |
| *Clathrocaspia gmelini* (Dyb.) | 262 | 162 | 33 | — | 457 |
| *C. pallasi* (Dyb.) | 5 | 17 | 396 | 3 | 421 |
| *Turricaspia spica* (Eichw.) | 222 | 4 | 56 | — | 282 |
| *T. spica* (Eichw.) var. *nova* Dyb. et. Gr. | — | 24 | — | — | 24 |
| *T. nossovi* n. sp. | 9 | — | 18 | — | 27 |
| *T. turricula* (Dyb.) | 20 | 44 | 14 | — | 78 |
| *T. grimmi* (Dyb.) | — | — | 46 | — | 46 |
| *T. bakuana* n. sp. | — | 7 | — | — | 7 |
| *T. dimidiata* (Eichw.) | — | 1 | 117 | 1 | 119 |
| *T. sieversi* (Cless.) | 77 | 2 | 1 | — | 80 |
| *T. pseudodimidiata* Dyb. et Gr. | 203 | 6 | 3 | — | 212 |
| *Planorbis (Gyraulus) eichwaldi* Grimm. | — | 2 | 38 | — | 40 |
| *P. (G.) eichwaldi* Grimm var. *dybowskii* nom. nov. | 3 | — | 19 | 2 | 24 |
| | 24122 | 1040 | 1517 | 16 | 26695 |

In the southern Caspian Sea three types of gastropod fauna may be distinguished. The first inhabits the shallower parts of the sea (4—20 m) and is characterized by a predominance of *Hydrobia*, *Caspia conus* (Eichw.). *Lithoglyphus* and *Theodoxus*. The second type is found at 20—80 m. It is far less abundant, but at 20—40 m differs little from the first type in respect to specific composition; significant changes begin to appear only below 40 m, where along with shallow water forms occur a great number of deep

water forms characteristic of the third type of fauna, inhabiting depths greater than 80 m.

The study of the horizontal distribution of species brought but few new data. Quantitatively the richest stations are those of the western coast, the stations of the eastern coast rank next, while those of the central parts of the sea are the poorest; but in respect to the number of species there is no appreciable difference. Thus, for instance, 24 forms and varieties were found at the western coast, 23 at the eastern, 18 between Baku and Krasnovodsk, and 21 between Lenkoran and Ogurchinsk island.

As regards the association of species, it may be noted that *Hydrobia*, *Zagrabica*, *Caspia conus* (Eichw.), *Clathrocaspia gmelini* (Dyb.) seem to agree well with one another. An increase in the number of shells of any of them is not followed by an appreciable decrease in the number of others. But a predominance of *Lithoglyphus* brings about a sharp decrease in the number of *Hydrobia* and *Clathrocaspia gmelini* (Dyb.), whereas the number of *Caspia conus* (Eichw.) and *Zagrabia* remains almost the same. *Hydrobia*, *Clathrocaspia gmelini* (Dyb.) and *Lithoglyphus* live among algae, and it seems that representatives of the latter genus compete with those of the two former.

It has also been observed that an increase in the number of *Theodoxus* was followed by a strong decrease of specimens of all other genera. However, this is another kind of competition. *Theodoxus* prefer a hard bottom, less favourable to the other gastropods of the Caspian Sea.

All these conclusions are of a preliminary nature, being based on a study of dead material (empty shells) which, though abundant, needs to be corroborated by further observations on leaving gastropods.

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*EXPERIMENTAL MORPHOLOGY*

# CONTRIBUTION TO THE STUDY OF THE RATE OF HEALING OF CUTANEOUS WOUNDS ON THE EXTREMITIES IN MICE AS DEPENDENT ON ENERVATION

By F. A. GUREVICH

*(Communicated by I. I. Schmalhausen, Member of the Academy, 18. II. 1941)*

The rôle played by the nervous system in regenerative processes has been rather thoroughly investigated for worms and amphibians ([1-3]). As for man and the mammals many points still remain obscure, and this notwithstanding the numerous studies which point to the important part played by the nervous system in the processes of regeneration and healing of wounds. This is why we found it expedient to report the results of some of our experiments and observations which had been suggested to us by B. P. Tokin.

All of them were made with white mice. The left hind extremity was shaven or carefully depilated, after which 0.5 $cm^3$ novocain were introduced subcutaneously within the region innervated by n. cutaneous fem. post. After the given region has ceased to respond to any stimulation, the skin was carefully pulled aside with the aid of a forceps, and an incision about 4—5 mm long was made along the main stem of nervus ischiadicus. Musculus gluteus maximus was next separated from the adjoining musculi by means of a forceps, fascia lata was ruptured, and a piece of nervus ischiadicus 3 mm long was excised well above the level of n. cutaneus fem. post. The wound was sutured with silk. Sterile conditions were observed but in rare cases. A total of 31 mice were operated upon. There were no cases of suppuration.

Our experiments fall into two series, one involving six mice, and another involving five.

1. Rate of Healing of Cutaneous Wounds on an Extremity in the Case of Enervation. Bits of skin, 14—13 $mm^2$ each, were excised by means of a sharp knife within the region innervated by n. cutan. fem. post. on both hind extremities. The left one was enervated, while the right one served as control. Observations on the rate of healing revealed the following picture. After bleeding has stopped, the edges of the wound begin to adhere. In some cases this takes place within the first 24 hours. Adhesion assumes a different course in the control extremity and in the enervated one. Thus, in 5 mice out of 6, the edges of the wound on the control extremity were found to adhere after 24 hours, which was never observed on the enervated extremity. After adhesion has taken place, clots of blood are formed between the edges of the wound. On the second

day epithelization sets in, beginning with the borders of the wound. In no case did we happen to observe on an enerved extremity adhesion of the wound edges within such a short period of time. In such cases the wound becomes covered with a scab within 24 hours; after a complete epithelization of the wound surface the scab falls off. On control extremities the wounds took about 4—5 days to heal completely, whereas on the enervated extremities healing took place in no less than 7—8 days.

2. Rate of Healing of Cutaneous Wounds on an Enervated Extremity under Local Anaesthesia. For the second run of experiments we took mice with their left hind extremity enervated; equal bits of skin, 14—13 mm² large, were excised on both hind legs. As the excision of a portion of n. ischiadicus brings about no full enervation, we smeared the wound on the enervated extremity with a 5% solution of dicain every 2—3 hours, which makes on the whole 25 times, our experiments having lasted from 27.VIII till 5.IX. In the meanwhile the control wound was smeared as many times with Ringer solution. While smearing, precautions were taken not to inflict any additional trauma. The process of healing observed in this run of experiments does not differ from that described for the first run. The results of our observations have shown that on the control extremity the regenerative process is completed within 5—6 days, whereas on an enervated extremity the same process takes twice as much time (10—11 days).

The experiments described above do not give us right to make any positive conclusions; the fact of a lower rate of healing of cutaneous wounds in the case of disturbances in the activity of the nervous system is, however, unquestionable. Our experiments do not give sufficient material to explain the mechanism of this phenomenon. It may, however, be assumed that healing of defects, as any other regenerative process, is a process of building of a strictly typical form. Its normal course must depend in the first line on the tissues and organs surrounding the wound. In the case of severe disturbances of normal functions and structures (blood supply and functions of the nervous system) the normal course of the form-building process, taking place in the tissues and organs adjoining the wound, should inevitably be disturbed.

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*EXPERIMENTAL MORPHOLOGY*

# HEALING OF THE WOUNDS AS INFLUENCED BY SURROUNDING TISSUES

By F. A. GUREVICH and V. E. CHECHETKIN

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The influence of a regenerative process taking place in a certain organ upon the organism as a whole has been proved experimentally. Aizupet (1) has shown that the regenerative process in *Rana temporaria* tadpoles influences the rate of metamorphosis and may severely disturb the normal course of ontogenesis. Upon thorough analysis, however, the problem appears to be still obscure and far from being solved definitely. Regeneration is, as it seems, a process local «par excellence». Viewed from this point, the rôle of the organism as a whole is restricted to creating the necessary background for a normal existence of cells on the wound surface and for a normal course of the processes, going on in tissues and organs adjacent to the wound. We are inclined, nevertheless, to consider the process of regeneration in the first place as a means for reproducing absolutely typical tissues, parts of organs, entire organs and systems of the latter. This form-building process of tissue restitution and development of organs seems to depend solely upon the nearest surroundings, i. e the «residual organ» (2). P. Weiss' experiments, as well as data by many other authors who were working on amphibia, provide convincing evidence as to this point of view: both the anterior and the posterior extremities of the newt were amputated, and the amputated posterior extremity was grafted to the wound surface of the removed anterior extremity. After this the grafted extremity was amputated so as to leave a disk-like residue. As a result of subsequent regeneration, a posterior rather than an anterior extremity was formed. It is clear that in this case the course of the formative processs was determined exclusively by the architectonics and by the functions of the small disk-like residue of the former posterior extremity (3). The aim of our experiment was to check these observations on mammals, though we were aware that surgical practice disposed of many a valuable materials bearing on this problem.

A circular groove *A* was cut in the skin and in the subcutaneous tissue on one side of a rat (Fig. 1). Then a circular incision *C* was made in the skin islet *B*. A similar circular incision *C* was made for control on the opposite side of the rat. Owing to the operation, the flap *B* proves to be isolated from the surrounding skin; it is in a state of semi-isolation from the organism, its normal nervous and vascular connexions being discontinued. If this is the case, then the wound *C* in the isolated region *B* must heal with a greater difficulty than an identical wound on the control side. This is true, of course,

only on the condition that regeneration, which is a typical form-building process, is not of a local character, but depends on the organism as a whole. If, on the contrary, one is to adopt the theory advanced by this laboratory, then no delay in the rate of regeneration should ensue.

In the first run of experiments 10 rats were operated upon, the result being 10 wounds $C$ 31. 4 mm² large on one side of each animal. The circular groove $A$ was 2 mm large. The islet $B$ had 2.3 cm in diameter. The wounds healed by secondary adhesion. In 6 rats the process of healing of both the control and experimental wounds was terminated simultaneously by the 7th day after the operation. In two rats the wounds on the control side healed on the 8th day, while those on the experimental side did so on the 7th day. In two other rats both wounds on the control and the experimental side healed on the 8th day. From this series of experiments it could be clearly seen that the healing of the wounds on the experimental side proceeded as successfully, as on the control one.

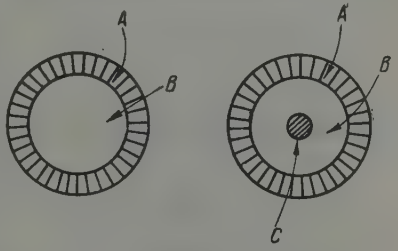


Fig. 1.
Consecutive stages of the operation.

The second run of experiments did not differ essentially from the first one. Our aim was to check our assumption on a more extensive material. Besides, in order to give a preliminary reply to the question as to how large an «organ residue» should be to secure a normal course of the regenerative process, we tried and reduced the size of the flap $B$ as much as possible. With this aim in view circular incisions $C$ 25.98 mm² have been made on both sides of 5 rats. The groove $A$ on the experimental side was made 2 mm wide. The width of the flap $B$ was but 4 mm.

In the course of these experiments in four rats both wounds, on the experimental and on the control side, healed at the same time (on the 7th day). In one rat the wound on the control side healed on the 6th day, while on the experimental side it healed one day later.

In the third run of experiments we made the flap $B$ half as small as that of the 2nd series, while the size of the wound $C$ was made twice as large. Thus, in 5 rats the circular incisions $C$ were made 50 mm² large. The groove $A$ was made 2 mm wide, so that the wound $C$ proved to be surrounded with a skin region but 2 mm wide. However, these experiments failed to yield the results expected, since in all the 5 rats the narrow skin flap began to necrose as soon as the next day.

The results of these 3 series of experiments go to confirm our idea of regeneration as a process, the typical course of which is dependent on the «organ residue», as well as on the surrounding tissues and organs.

The present work has been suggested by Prof. B. P. Tokin.

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*HYDROBIOLOGY*

# BIOLOGICAL SEASONS IN THE PLANKTON OF DIFFERENT SEAS

By B. G. BOGOROV

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Seasonal phenomena in marine biology are of general occurrence, although their manifestation varies as to the latitude. The definite periods of seasonal changes in the biology of the sea, the so-called biological seasons ([1], [2]), are directly related to hydrological seasons, while the latter are related to climatic seasons.

Biological seasons are manifest in the life of all the habitants of the sea, from bacteria to whales. All the fishing activities of men are being closely adapted to seasonal phenomena. The best known of them are those taking place in the life of the plankton. However, investigators do usually content themselves by working out plankton calendars (Steuer 1930, Zernov 1934), though, obviously, no clear understanding of biological phenomena can be gained by merely referring certain amounts, or certain specific compositions of plankton to definite months. The qualitative and quantitative variations in the plankton must be traced not after formal calendars, but by the course of the biological processes which determine the succession of biological seasons. To obtain a full characteristic of the biological seasons, different phenomena of the biology of organisms should be considered, as well as variations in the environmental conditions (table *).

An analysis of the table shows that the course of biological seasons and the transition from one seasonal state, or its stage, to another is determined, on the one hand, by internal biological causes (quantitative and qualitative changes in the plankton, physiological state of organisms, especially age, reproductive power, etc.), and on the other hand, by the development of the external conditions of the regime prevailing in the given sea, such as: intensity of illumination, scattering of the ice, quantity of nutrient salts, water temperature, discharge of fresh water, etc.

Let us follow the course of seasonal variations of the plankton, i. e. of biological seasons in different latitudes from the North Pole to the equator. For this purpose, we may use the works of numerous authors (the most important of them are cited in our list of references). Our diagrams (Figs. 1 and 2) are but schematic. Every given sea may of course somewhat deviate from the scheme obtained. Thus for instance, the second maximum may be not smaller,

* This table must be regarded as an approximative scheme, which gives some general features of different biological seasons, as it is impossible to give a fuller characteristic in so short an article.

Characteristics of Biological Seasons for the Plankton of Different Seas (Superficial, Photosynthetic Layer)

| Biological seasons | | Spring | Summer | Autumn | Winter |
|---|---|---|---|---|---|
| Amount of plankton. Relation between vegetable and animal plankton | | Yearly maximum. Particular abundance of phytoplankton. Yearly maximum of «bloom». Biomass of phytoplankton by many times exceeding the biomass of zooplankton | Less plankton than in spring. Abundance of zooplankton — yearly maximum. Its biomass equal to or exceeding that of phytoplankton | Sharp decrease of plankton in polar seas, and an increase in southern seas. Second maximum of phytoplankton. Its biomass greater than that of zooplankton | General minimum of plankton. The biomass of zooplankton greater than that of phytoplankton |
| Qualitative changes of the plankton, age composition, generations, growth rate, etc. | Zooplankton | Many short-lived species. Great masses of eggs, nauplii, larvae of benthic animals. One generation of copepods. The organisms of the spring generations are the largest; the generation is very numerous. Many deep water animals ascend to superficial layers for breeding. Plankton-eating fishes start feeding; they keep on the edges of the «bloom» patches | Appearance of warm-water species. Many larvae of benthic animals. In seas with long summer seasons the specific composition varies but little. In southern seas copepods give 2, 3, 4 generations. Maximum growth rate. The size of organisms decreases. Slow developing organisms «grow older». Plankton-eating fishes feed intensely; they keep among aggregations of copepods and other small crustaceans, often rather scattered than in schools | Predominance of warm-water organisms (especially in the beginning). Larvae and young organisms disappear. A last generation of copepods usually with the smallest size of organisms. Long-lived species continue to «grow older». Plankton-eating fishes feed less intensely or stop feeding. They keep in great schools | Warm-water species and young organisms disappear. Survive adult and hibernating stages. Plankton-eating fishes almost cease to feed; they keep in great schools |
| | Phytoplankton | Mass development of species needing for their development high concentrations of biogenic elements (nutrient salts) | Species needing no high concentrations of nutrient salts become predominant | Peridineae algae become most important along with diatomeae | Algae are in a depressed condition, often in stages of rest or hibernating cells |
| Chemical, hydrological and climatic features | | In the beginning of this period biogenic elements (nutrient salts) occur in great amounts. Superficial waters usually are strongly freshened. Low water temperature. In polar seas appearance of open ice and great openings. Intensity of illumination increases. Air begins to warm | Minimum amount of biogenic elements. Increase in salinity and temperature of water. In polar seas freshening of water due to intense melting of ice. Maximum intensity of illumination and air temperature | More biogenic elements than in summer. Water $t°$ maximum in the beginning drops toward the end of fall. Water salinity lower than in summer. Intensity of illumination decreases. Air grows colder. Atmospheric precipitations | Amount of biogenic elements increases, reaching its maximum toward the end of the winter. Lowest level of air and water $t°$. Appearance of ice. Maximum salinity in the Antarctic and freshening of water in southern seas. Minimum intensity of light |

but even greater than the first. However, a certain degree of conventionality in plotting the general curve of the annual variations of the plankton could not possibly be avoided. Our plots are based on materials obtained by different methods (determinations of biomass, number of organisms, indications to abundance, etc.). Thus, it was impossible to work out a vertical scale for the curves of different seas. However, these data are sufficient to determine the periods of mass development of organisms, duration of biological seasons and their distribution within the year.

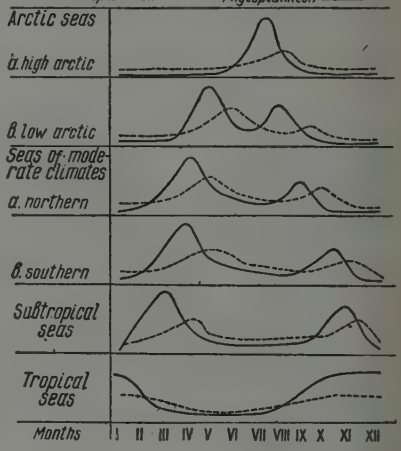

Fig. 1.

An analysis of the plots permits to establish the following regularities:

1) Biological seasons in seas of different latitudes correspond to different months of the calendar. Thus, for instance, in tropical seas the biological spring coincides with the winter months, whereas in the Polar basin it sets in in August.

2) The duration of biological seasons (especially of the biological winter and summer) in different latitudes is subject to great variations. In tropical seas the biological summer lasts 7 months, while in polar seas it is only one month long. The winter, which is scarcely expressed in the tropics, lasts 9—10 months in polar seas.

3) In different seas during the year there may occur one or two maxima of development and of abundance of phytoplankton and hence of zooplankton.

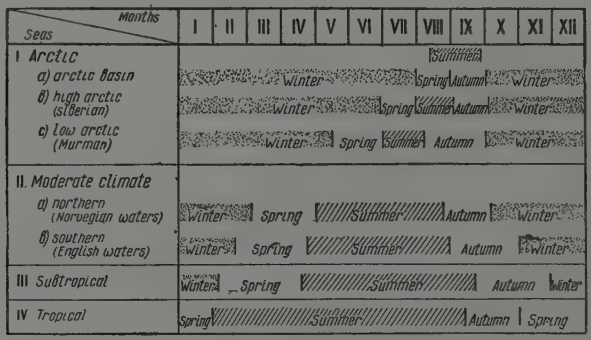

Fig. 2. Duration of biological seasons.

4) The time intervals between the main maxima vary with latitude, decreasing towards the North.

5) The period of growth decreases from 12 months in tropical and temperate seas to 2—3 months in polar seas.

6) The plankton of each sea is subject to such seasonal variations, that its quantities in different seas can be compared only after a material obtained within the same biological seasons.

All these regularities of the course of biological seasons of the plankton are related, mainly, to the seasonal changes of temperature and light. Temperature is a particularly important factor in tropical, subtropical and moderate seas, whereas in polar seas solar illumination is the main limiting factor, and the peculiarities of biological seasons in high-arctic seas are related to the penetration of light in the water.

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*PARASITOLOGY*

## A NEW VECTOR OF THE TICK RELAPSING FEVER—*ORNITHODORUS NEREENSIS* PAVL. IN TURKMENIA

By E. N. PAVLOVSKY, Member of the Academy

In the course of the investigations made in 1931 by the Karakalin and Kzyl-Atrek expeditions (south-western part of the Turkmenian SSR) we found three species of ticks belonging to the genus *Ornithodorus*. *Orn. papillipes* was found near Kara-Kala in a grotto-like cave. Shallow burrows and crevices of clay and stone huts, as well as upturned stones, often harboured ticks which have provisionally been referred by me to a new species, *Orn. nereensis* in litt. I photographed the biotopes which were their habitat and published these photoes in 1934 (biotopes *E*, *G*, *Z*, *I*, *K*, figs. 4, 6, 7). Turtle holes were also found to contain *Orn. tartakovskyi*, which has been met in Kzyl-Atrek and Bayat-Khadji as well. Thus, this Turkmenia district is inhabited by three *Ornithodorus* species. Testing of the material collected in 1931 for its spontaneous infection with spirochaetes of the tick relapsing fever gave negative results. The material collected in 1931 being rather scanty, a group consisting of scientific worker M. W. Pospelova-Strom and two assistants were sent from my laboratory (Department of Parasitology of the Institute of Experimental Medicine) in 1936 to collect tick materials. A rather abundant material was collected which could be utilized both for the faunistic investigation of M. Pospelova-Strom, and for my present paper on the possible rôle of *Ornithodorus* found in this region as vector of the tick relapsing fever. Nothing but verbal statements concerning rare singular cases of tick relapsing fever could be gathered in Turkmenia. Nevertheless an analysis of the relating materials is most interesting. The more so as Kzyl-Atrek stretches like a wedge far southwards, and the Iran border line is known to be unsecure with regard to tick relapsing fever. In particular, Astrabad has been long known to be a nidus of relapsing fever; data concerning Russian soldiers in Ardebil enabled Djunkovski to describe in 1912 *Spirochaeta persica*. Therefore the inspection of the adjacent regions of Turkmenia becomes of special interest for the knowledge of geographical distribution of the tick relapsing fever.

Control feeding of the ticks collected in 1936 was carried out on guinea-pigs and white mice. *Ornithodorus papillipes* originating from a cave in Arpaklen tested 25. X. 1936 and 21. X. 1938 gave negative results, as well as ticks from the Kizyl-Imam cave (tested 10. VIII. 1937, 25. II. 1939 and 11 . VII. 1940); the batch of ticks collected in Kizyl-Imam was regularly multiplied under laboratory conditions during 1938, 1939 and 1940. A few ticks which on 23. VII. 1938 had been fed on a bat by M. Pospelova-Strom showed nega-

tive results on the 3rd of August 1937, whereas on the 10th of July 1940 they were able to cause spirochaetosis in guinea-pigs. Bats may be carriers of spirochaetes causing tick relapsing fever; in virtue of this it could not be decided in the present case, whether the ticks tested had been spontaneously infected with spirochaetes, or whether they obtained the latter when feeding on bats?

Ticks from a porcupine burrow in Nere (vicinity of Kara-Kala, collected 15. VII. 1936) infected guinea-pigs on 13. IX. 1936, 21. II. 1939 and 13. III. 1940; the same ticks multiplied yearly under laboratory conditions, since 1937 till 1940. The spontaneous infection of ticks *Orn. nereensis* was tested on white mice. Yields from holes in the duval at Nere dating 15. VII. 1936 gave twice a positive effect as to their spontaneous infectivity; spirochaetes were found in mice on 28. IX. 1936 and 28. VII. 1937, 3—4 ticks feeding simultaneously on one mouse. Other batches of ticks from burrows in Kara-Kala and found under stones on mountain slopes showed infectivity but once; subsequent feedings were resultless, this being evidently due to dying out of infected ticks. Thus, the fact of spontaneous infection of *Orn. nereensis* with spirochaetes and the ability to transmit the latter while sucking blood has been established.

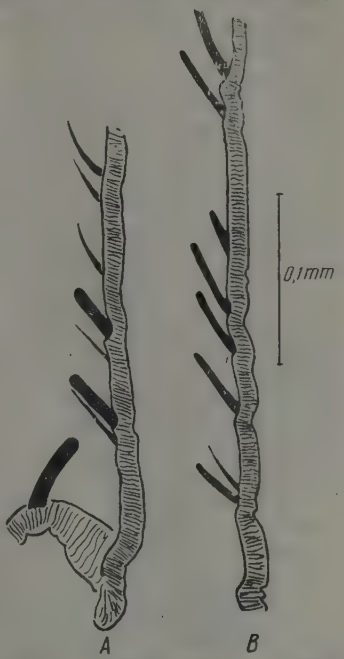

Fig. 1. Ventral margin of tarsus: *A*—of first, *B*—of fourth leg in *Ornithodorus nereensis* Pavl.

The following examples show the course of spirochaetosis in mice, bit by *Orn. nereensis*. Mouse No. 75 was infected by feeding on it 3 ticks on 28. VII. 1937. Investigation of blood for the spirochaete gave the following results:

| August | 1 | 3 | 5 | 7 | 9 | 10 | 13 | 15 | 17 | 19 | 21 |
|---|---|---|---|---|---|---|---|---|---|---|---|
| | (+)— | ++ | + | ++ | + | — | ++ ++ | — | — | — | — |

Mouse No. 81 infected on 31. VII. 1937 by means of ticks from the cave of the aul Dickeldash valley; investigation for spirochaete presence in blood:

| August | 4 | 5 | 7 | 9 | 11 | 13 | 15 | 17 | 19 |
|---|---|---|---|---|---|---|---|---|---|
| | — | — | (+)— | ++ | ++ ++ | ++ ++ | — | + | — |

The spirochaetosis in mice is of very short duration. Ticks *Orn. nereensis* may transmit spirochaetes to man (treatment of progressive paralysis with experimental tick relapsing fever, transmitted by *Orn. nereensis* bites).

Thus, two out of three *Ornithodorus* species of the Kara-Kala region fauna were found to be spontaneously infected with spirochaetes of tick relapsing fever. The fact that *Orn. papillipes* acts as a vector of this disease has been established long ago. *Orn. nereensis* presents a double example, both of a new tick species and a new vector. The general habitus of *Orn. nereensis* approaches that of *Orn. erraticus* L. (syn. *O. marrocanus* Velu.), common in North Africa and Spain, though differing from the latter by a series of distinctions: the tarsus of the 1st pair of legs of *O. nereensis* bears besides sharp hairs

1—3 broad, short and blunt spurs (rod-like hairs) along the posterior ventral margin of the leg. The II and III segments of the palps of the male of *O. nereensis* bear rod-like hairs, while in the case of the female the palp hairs are sharp. The tarsus of the IV leg bears along its ventral margin short rod-like hairs. The spiracle of *O. nereensis* possesses a narrow free hoop, from which rods of irregular shape proceed on to the centre along the radii.

Diagnosis of the Species. Body oblong; chitin covered with tubercles, shagreen-like, grey in live ticks; the chitinous tubercles rather large, hill-like with rounded summit and ribbed folds along the slopes. Along the sides of the capitulum there are «cheeks» consisting of separate digitate sclerites. Chitinous integument of males above the capitulum, between the cheeks and up to the rostrum, bears singular chitinous hairs (ranging from 10—11). The anterior ones are blunt rods, the rest are sharp. Females possess more numerous hairs than the males. The dorsal margin of the tarsus of I leg is uneven, forming two broad waves; besides common ones, the ventral margin bears 1—3 blunt, short rod-like hairs; the femur of I leg is supplied with blunt rod-like hairs with rounded ends. Dorsal margin of tarsus of IV leg is even; near the end it becomes bent forming an obtuse angle with the basis of ungues; ventral margin of tarsus of IV leg bears blunt rods and sharp hairs. Genital aperture of male lies on the level of the line between coxae of I—II legs, whereas in the female it is on the level of II coxae. The spiracle is on an oblong protrusion. The II—III segments of the palps of the male bear rod-like hairs; all the hairs on the palps of the male are sharp. Spiracle with few rods disposed along the radii. Length of female 3.9—4.2 mm; length of male 2.8 mm; width of female 2.5 mm, that of the male 1.6 mm. Pests of mammals, reptilia and birds living in burrows.

Additional investigation is required to decide whether *O. tartakovskyi* from Kzyl-Atrek and Kara-Kala is a carrier of spirochaetes, since the negative results of the investigation cannot be considered decisive in view of the scanty material collected. It will be no surprise, if *O. tartakovskyi* is found to show spontaneous infection with spirochaetes, since the related species from Mangyshlak was found to be the carrier and vector of spirochaetes (Pavlovsky, 1939), while M. Pospelova-Strom (1941) reported the same of *O. tartakovskyi* from the vicinity of Ashkhabad. The Kara-Kala region is so far the only place where *Orn. nereensis* is being found in Middle Asia. A respective investigation of North Iran would be of interest.

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*PARASITOLOGY*

# ON SOME INTERRELATIONS BETWEEN TICKS AND INSECTS

By E. N. PAVLOVSKY, Member of the Academy

As it is known, the *Ixodidae* ticks are ectoparasites on the vertebrates—mammals, birds, reptilia, and less frequently, amphibia; this circumstance is based on the fact that the ticks are adapted to feeding on blood; it should be noted that, besides blood, ticks suck in pathological products from the skin region of the host, altered by the action of the saliva of the tick. Judging by pathological changes caused by its saliva, the effect of local action of a sucking tick is very great. The said effect increases still in the case when numerous ticks feed on a limited skin area.

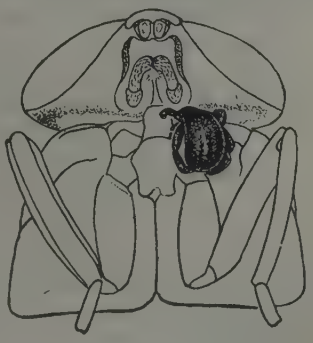

Fig. 1.

Insects do not count among natural hosts of ticks belonging to the *Ixodidae*. There have been reported but isolated instances of ticks attacking their own like; thus, cannibalism has been observed in the case of *Ornithodorus verrucosus*. A. Scrynnik (4) described a case when two males were found fixed on the ventral surface of a female. The attachment of a male *Boophilus calcaratus* to a horse-fly *Tabanus tergestinus* Egg. seems to be a unique case. The tick became attached from the ventral side to the neck of the horse-fly preventing movements of the left anterior extremity (Fig. 1). The horse-fly was caught on a cow in Muchrovani (Kakhetia) on the 1st of August 1931; cows were infested considerably with ticks *Boophilus calcaratus* (the material was obtained by courtesy of N. Matikashwili). *B. calcaratus* is a single host tick; males hatch from nymphae on the cattle and copulate in situ with females; the males often attach themselves to the skin of the host near a female sucking blood. The males themselves do not suck.

In the present case the male could attach itself to the horse-fly only while the latter was busy sucking blood, which does not take long. Judging by the spot of the horse-fly's body to which the tick became attached, the *Boophilus* male even trying to do his best was unable to penetrate farther than the haemolymph of the insect, but had no chances to reach the blood sucked in by the insect, since it does not stop in the oesophagus but flows rapidly through it into the stomach of the horse-fly. Evidently, in this case the male tick, owing to alteration of its instinct, instead of fixing itself on the skin of the cow became erroneously attached to the sucking horse-fly.

Cases are known when some *Arachnioidea* and insects attached themselves to insects; for example, pseudoscorpions are often found attached to flies, Tyroglyphidae ticks—to various bees, Trichodectes to pupipara flies, etc. The attached individuals behave as passangers on their temporary host; with its aid they can reach such surroundings which are ecologically favourable to their further existence and breeding. Such a passive way of spreading through transporting hosts (zoochoria), to whose body surface ticks and insects attach themselves, is known under the term «forezia».

The case under description cannot be referred to the forezia category, since the ticks *Boophilus* remain throughout their life cycle on the cattle, infesting the latter in the larval stage; therefore, the finding of their host-animal is entirely the business of the larvae; adult ticks do not seek their hosts, since males and females hatch from nymphae on the skin of the cattle, where they copulate and become attached for feeding.

Evidently, the case of attachment of a *Boophilus calcaratus* male to a horse-fly should be explained by a peculiar case of «misled instinct» of attachment. It should be noted that up to now no cases of forezia through horse-flies have been described.

Another case of a certain interest concerns the pasture tick *Hyalomma* and the bed-bug. Among the material collected in Southern Tadjikistan I found one individual of *Hyalomma detritum* whose leg bore a bed-bug egg attached to it. As noticed by me before (1), the adults of *Hyalomma* molt from nymphs under the cover of manure bricks (fuel storage), which, as it is the custom in Tadjikistan kishlacks (villages), are kept outside and inside clay walls of cattle stalls. Owing to this, *Hyalomma* may come into close contact with bed-bugs which are frequent inhabitants of henhouses. The case in question is explained by such a temporary cohabitation of the bed-bug and the tick *Hyalomma*, under conditions created by man.

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*PARASITOLOGY*

# INFLUENCE OF PARASITIC WORMS ON THE SECRETORY FUNCTION OF THE STOMACH AND SMALL INTESTINE IN DOG

By TH. TH. TALYSIN

*(Communicated by E. N. Pavlovsky, Member of the Academy, 10. I. 1941)*

Notwithstanding the fact repeatedly mentioned in the scientific literature ([1], [2], [4]) that the secretory function of humans and animals infested with parasitic worms becomes altered, there exist no direct experimental proofs bearing on this subject; with the aim to fill up this gap we undertook the following investigation.

I. Action of Parasitic Worms on the Gastric Secretion. Examination of the gastric juice was made on a dog according to Pavlov's method of isolated pouch. The dog «Arto», a male, weight 24.5 kg, was fed on common mixed food. On the day of the experiment the animal was given only 200 g of minced meat. A total of 73 experiments were carried out at 2 days' intervals; each experiment lasted 5 hours.

Observations on the dog were carried out according to the following scheme: length of latent period and volume of juice secreted every 15 min. Along with this quantitative control, the following items were determined: general acidity, free HCl in %, and digestive power of juice after Mett. The acidity was determined by means of $^1/_{50}$ *N* NaOH solution, 1% alcohol solution of phenolphthalein and a solution of dimethylamidoazobenzene used as indicators. Analysis of faeces showed that the animal under experiment was free of worm parasites.

Having determined the normal volume of the secretion of gastric juice, which in the average for 5 hours equalled 44.4 cm³, we began to feed the animal with worm material (after the method of E. N. Pavlovsky). In some of the experiments dried parasitic worms, ground to thin powder, were emulgated in saline and fed to the dog; while in other experiments we administered emulsions prepared with fresh worms obtained either from a helminthological point, or the slaughter house. In the course of all the experiments we have tested: suspensions of dried *Taeniarhynchus saginatus*, *Diphyllobotrium latum*, *Ascaris lumbricoides* (females) and emulsions of fresh parasites *Acanthocephalus hyrundinaceus* (females) and *Ascaris lumbricoides* (males, as well as females) which were added to the meat fed to the dog. Intervals ranging from 4 to 14 days between two consecutive feedings were made in order to observe the after-effect.

A feature common to all the experiments was the pronounced inhibiting effect of the substances introduced on the secretion of the canine stomach.

*T. saginatus* yielded the highest effect reducing the mean volume of juice secreted within 5 hours to 30.2 cm³, which constitutes 34.5% of the normal secretion within the same period of time. The reduction in the volume of juice secreted began on the 1st day after the dog was given 15 g of dried strobilae of the tapeworm *T. saginatus*. Inhibition of the secretory activity was observed not only when the dog was fed on parasites, but during the following period, as well. So in the case of *T. saginatus*, inhibited secretion was observed by us in the course of 27 days (Fig. 1), which points to the fact that the inhibitory effect persisted after the helminthic substances have been absorbed. In the course of the experiment the gross volume of juice never reached normal values. It should be noted besides that inhibited secretion could be observed mainly at the second—humoral phase of juice secretion, while

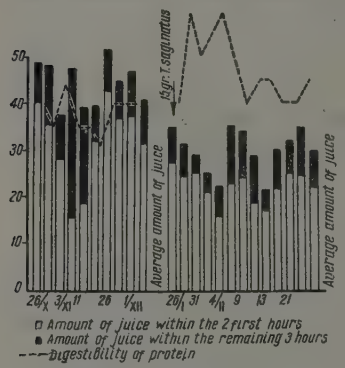


Fig. 1. Effect of *T. saginatus* on the pouch isolated according to Pavlov's method: norm (left), effect of 15 g *T. saginatus* (right).

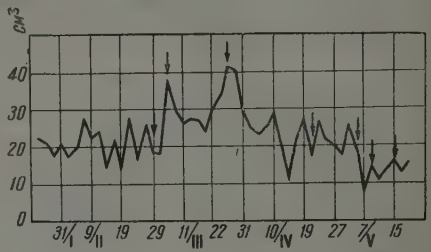


Fig. 2. ↓—Days when the dog with an intestine loop isolated after Tiri was given 15 g *T. saginatus*.

during the first two hours, corresponding to the reflex phase, the secretion exceeded the norm by 26.5%. Repeated feedings went to increase the inhibitory effect. Unlike to this, the digestive power of the juice showed in most experiments a slight increase as compared to the norm. In controls it reached in the mean 3.7 mm, while after feeding with *T. saginatus* it amounted to 4.7 (after Mett). The acidity, both general and free HCl tended to decrease after feeding with *T. saginatus*, *D. latum* and *A. lumbricoides*.

The following is a record of the control experiment and that of feeding the dog with *T. saginatus*.

| Hours of experiment | Experiment 25; 24 Jan. 1940; control | | | | | Experiment 26; 26 Jan. 1940; day of feeding | | | | |
|---|---|---|---|---|---|---|---|---|---|---|
| | Volume of juice in cm³ | Acidity in % | | | | Volume of juice in cm³ | Acidity in % | | | |
| | | general | | | free HCl | | | general | | free HCl |
| 1 | 16.4 | 0.496 | | | 0.569 | 13.5 | 0.496 | | | 0.584 |
| 2 | 13.9 | | | | | 11.8 | | | | |
| 3 | 9.6 | 0.511 | 0.525 | | 0.598 | 6.3 | 0.477 | 0.477 | | 0.555 |
| 4 | 3.9 | | | 0.598 | | 2.3 | | | 0.584 | |
| 5 | 2.3 | | | | | 1.0 | | | | |
| Total | 46.1 | | | | | 34.9 | | | | |

*Diphyllobothrium latum*. In the course of experiments when a suspension containing 5 g of dry strobilae was given, the volume of juice secreted became reduced in the 6th experiment to 30.7 cm³ (norm—35.0 cm³), i. e. by 12.3%. 20 g emulsion prepared with strobilae of *D. latum* for 5 experiments gave a mean decrease by 12.8%. Thus, there was inhibition of juice secretion in response to treatment with *D. latum*, though a less pronounced one than in the case of *T. saginatus*.

*Ascaris lumbricoides*. 2 g of suspension of dry female *Ascaris* bodies during 4 experiments caused a reduction in juice secretion; additional feeding with emulsion containing one emulgated fresh female of *Ascaris* weighing 9.8 g increased the inhibition: mean volume 32.5; i. e. a decrease equalled 4.4% (norm —33.9 cm³); a comparatively higher effect was produced by emulsions of 5 males, weighing 7 g; this brought about a decrease up to 31.5 cm³ (norm—33.7 cm³), i. e. by 6.6%.

*Acanthocephalus hyrundinaceus*. Emulsions of fresh entire female worms weighing 7 g caused a reduction in the volume of juice: mean value—40.5 (norm—44.1), which made up a difference equal to 8.7%.

Summing up the data obtained, the fact of inhibition of gastric secretion under the influence of feeding the dog with helminthic substances becomes quite evident. The inhibitory effect is also corroborated by the fact, that the secretory power of the dog's stomach regained its norm as soon as we stopped our experiments.

The lengthy after-effect (Fig. 1) resulting from a prolonged feeding with helminthic substances and expressed by a more or less pronounced reduction of the volume of gastric juice secreted and sometimes by a decrease of its acidity, is a most characteristic feature, as well. An increase in the digestive power of the juice was observed by us after feedings with *T. saginatus* only.

II. Effect of Parasitic Worms on the Intestine Secretion. The experiments were carried out on a dog with an intestine loop isolated after Tiri. Dog named «Dick», male, weight 17.5 kg, fed with common mixed food; during the experiments was in a good condition. The juice was obtained by means of a mechanical irritant (rubber drainage tube). The experiments were carried out every other day; length of each experiment—6 hours. Besides registering the volume of juice secreted, we tested it for amylase after Wohlgemuth. Analysis of faeces showed that the animal was not infested with intestinal parasites.

Suspensions of dried and ground strobilae of *Taeniarhynchus saginatus* were tested. Helminthic substance was diluted with saline, added to 100 g of minced meet, and fed to the dog.

Control experiments were carried out beforehand on the same dog, it was given same amounts of pure meat; juice secretion remained without changes. Two consecutive feedings with helminthic substances were separated by a 2—8 day interval. A total of 50 experiments were carried out.

Normal secretion (background) ranged from 14 to 27 cm³ within 6 hours. Increase and decrease of juice secretion were separated by an interval of 12 days and showed a rather regular character. The mean volume of juice secreted equalled 20 cm³ within 6 hours (15 observations). The results of the experiments may be expressed by the curve in Fig. 2.

The first experiments gave a sharp increase in the secret produced both on the day of feeding with *T. saginatus* and during the subsequent period. Repeated feedings caused an increase in the secretion. In separate cases the volume of juice reached 41.1 cm³, which exceeded the norm twice (Fig. 2). The next experiments brought about a reverse effect causing a decrease of secretion, this inhibition being of a lasting character. The minimum juice secretion was 7.9 cm³, i. e. almost $^1/_3$ of the norm. This could be observed

on the 2nd day following feeding. No changes whatsoever in the amylase contents were stated.

The activity of the glandular apparatus of the intestine usually becomes excited in response to local irritation. From literature it is known that various food components (except fats) do not affect the amount of intestinal juice secreted, when administered but once (Savich and others). As stated above, our control experiments with meat rations did not cause any changes in secretion either.

Experiments which involved feeding to the dog 100 g meat containing *T. saginatus* showed changes in juice secretion both on the day of feeding and during the following period. This indicates that the strobilae of *T. saginatus* contain some substances producing an unmistakable effect upon the function of the intestinal glands.

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